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SOUND AMPLIFICATION AND PUBLIC ADDRESS

by J. DE BOER.

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In both the open air and auditoria with defective acoustics the intelligibility can be improved by means of electrical amplifying equipment. The location of loudspeakers and microphone must be given careful thought in order to avoid undesirable conditions arising in addition to the intensification of the sound, such as an amplification of disturbing noises and an increase in the reverberation time. In many cases the acoustics can only be improved by means of an electrical amplifying equipment using microphones and loudspeakers with pronounced directional characteristics.

The introduction of suitable electrical amplifying equipment during the last few years has made it possible for the spoken word and music to be clearly heard at remote points where previously sound was inaudible owing to insufficient sound intensity. An electrical equipment for amplifying the sound waves and thus increasing the range of audibility and intelligibility consists essentially of three components:

1) The microphone, which converts the sound vibrations into electrical alternating voltages, embracing exactly the same frequencies as the sound vibrations themselves and each with an amplitude proportional to the amplitude of the initial atmospheric vibration;
2) the amplifier, which amplifies the alternating voltages in the requisite ratio, and
3) the loudspeaker which again converts the amplified alternating voltages into sound vibrations.

The applications of sound-amplifying equipment can be conveniently subdivided into two main groups:

A) For use where the speaker and audience are in the same auditorium, but the speaker's voice is not sufficiently intelligible to the whole audience without electrical amplification. In this case the electrical equipment serves in some measure to improve the acoustics of the auditorium. Sound amplification in the open air also comes under this group.
B) For use where the speaker is not in the same room as his audience, so that the electrical equipment acts as the direct source of sound in the auditorium.

In the present article we shall discuss mainly the first group of applications, *viz.*, the improvement of acoustic conditions and particularly the intelligibility of speech by means of loudspeakers.

When is sound amplification required?

An improvement of the acoustics by electrical amplifying equipment is required when satisfactory acoustic conditions cannot be achieved by other means, such as a suitable acoustic design of the auditorium or a correct choice of the wall and floor coverings, or where these measures cannot be taken for practical reasons, *e.g.* in an auditorium of too great a volume or in the open air. The problem can be conveniently discussed by a more detailed consideration of the following aspects:

1) The acoustics of an auditorium are satisfactory although its volume is too large, with the result that the natural sound intensity is insufficient for clear intelligibility or the source of sound is too feeble.
2) The geometrical shape of the auditorium does not give a uniform distribution of sound.
3) The auditorium is planned on the correct acoustic lines, but its time of reverberation is too high.
4) The audience is in the open air.

Sound amplification in an auditorium with good acoustics

If the auditorium is of the correct geometrical form so that the sound is uniformly distributed, and it has also a reasonable time of reverberation, the intelligibility can still be unsatisfactory if,

owing to the size of the auditorium or disturbing extraneous noises, the mean sound intensity is too low compared with the extraneous noise. By increasing the emission of sound from the source, satisfactory and adequate intelligibility would be obtained in this case.

We must examine how far it is possible to produce a more powerful source of sound by means of electrical amplifying equipment, assuming that the microphone is located in the centre of the auditorium in which the sound is uniformly distributed, and that the loudspeaker is placed close to the speaker.

It is first necessary to decide what effect an amplifying equipment of this type has on the acoustic characteristics of the auditorium. If the loudspeaker has no pronounced directional radiation, the sound distribution is not altered by the provision of an amplifying equipment, which indeed is quite unnecessary if the auditorium has been planned on efficient acoustic lines. But if an amplifying equipment is installed in an auditorium of this type the reverberation time will be automatically altered, and it will indeed be increased in all cases, as may be gathered from the following considerations: The sound intensity in an auditorium is, according to Sabine's theory, independent of the location of the sound source, and in the case in point with an auditorium in which the sound is uniformly distributed this theory naturally applies. The acoustics of the auditorium will therefore not be altered if the amplifying equipment is replaced by a surface which absorbs as much sound as impinges on the microphone and also radiates as much sound as the loudspeaker. This surface will thus have a negative absorption factor. The magnitude of this coefficient depends on the degree of sound amplification, the conditions here being exactly the same as when sound is amplified by reflection. As the reflected sound increases in intensity, the time of reverberation also increases. In fact, with an amplifying equipment the total absorption of sound by the walls of the auditorium can be reduced to zero. The sound waves will then no longer decay; this phenomenon is termed acoustic reaction.

If, as we have assumed, the sound intensity is still inadequate when the correct time of reverberation has been obtained, it is apparent from the above arguments that the installation of amplifying equipment will not result in any improvement. It is fortunate that in practice conditions are not as unfavourable as this. We have assumed that the sound intensity at the place of the microphone is equal to the mean intensity of sound in the auditorium. If the microphone is placed sufficiently close to the speaker, the intensity of sound reaching the microphone can be made much greater than the average sound intensity in the auditorium, with the result that the gain of the amplifier and hence also the negative absorption factor of the equipment can be reduced. With the same mean sound intensity the time of reverberation can again be slightly reduced by adopting this method, while at the same time the extraneous noises in the auditorium are less amplified than the speaker's voice.

A further improvement can be achieved by using a microphone with a directional characteristic, *i.e.* a microphone with a powerful response to sound waves travelling from the direction of the speaker, while its average response (the mean response calculated for all directions) is much smaller. The average response is, however, a measure of the negative absorption coefficient, as reverberant sound impinges on the microphone uniformly from all directions.

Directional Characteristics of Microphones

A microphone, which responds to the acoustic pressure, has the same response for all sound waves impinging on it irrespective of direction, if its dimensions are smaller than or of the same order of magnitude as the wavelength of the sound pulses. A pressure-gradient microphone possesses different response characteristics and reacts to the periodic difference in pressure between two adjoining points, for instance on the two sides of a diaphragm. A sound pulse whose direction of propagation is perpendicular to the line joining these two points produces no difference in pressure between them, and the microphone will therefore not respond to this sound pulse. If the direction of propagation makes an angle φ with this line the microphone will respond only to the component of the differential pressure along this line, the induced potential being thus proportional to $\cos \varphi$. In a directional diagram (*fig. 1*), the response of a microphone with these directional characteristics is represented by two circles making contact at the origin.

When using a combination of a pressure and a pressure gradient microphone, both having the same maximum response, the induced potential is proportional to: $^1/_2\ (1 + \cos \varphi)$. This directional effect is represented by a cardioid, which is also shown in fig. 1.

According to definition, the power response of a microphone is proportional to the square of the potential induced in it by a given acoustic field. If an average is struck of the response from all directions in space, it is found that the average response values

in the three directional diagrams discussed above are in a ratio of $1 : {}^1/_3 : {}^1/_3$ [1]), while the maximum response (at $\cos \varphi = 0$) is the same in all three cases. With the two last-mentioned types of microphone the negative absorption coefficient is thus three times smaller than in the pressure microphone, if the sound amplification is the same in all cases.

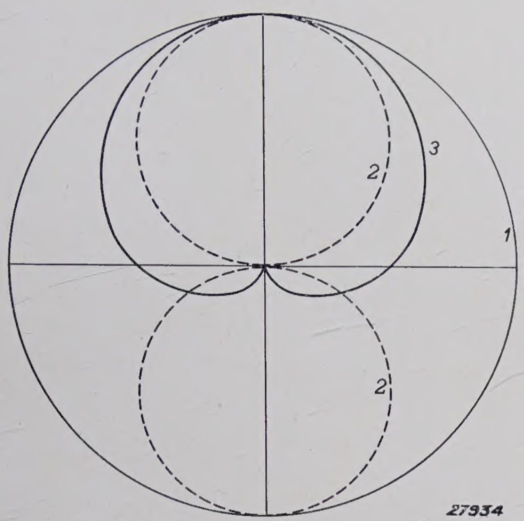

Fig. 1. Directional diagram of microphone characteristics, with the potential at the microphone at a given pressure amplitude plotted as a function of the direction of the sound pulse. *1.* Pressure microphone; *2.* pressure-gradient microphone; *3.* combined microphone (${}^1/_2$ pressure microphone + ${}^1/_2$ pressure-gradient microphone).

Position of loudspeakers

If the auditorium is designed such that the sound is uniformly distributed over it, the loudspeakers can be located at any convenient points. But since the ear can always detect the source of sound in spite of the equalisation of sound distribution produced by reflection, it is desirable to place the sound projectors close to the speaker. The perception of the direction from which a sound pulse is radiated is mainly due to the position of the ear relative to the direction from which the sound waves first impinge on it, *i.e.* the direction of propagation of the direct waves. If the wavelength of the sound pulses is of the same order of magnitude or smaller than the dimensions of the head (*i.e.* in the case of notes with a frequency above about 300 c/s), the sound intensity impressed on the two ears is not the same. The difference in intensity which depends on the direction of the incident sound is a measure of the directional impression. The time interval elapsing between the impression of sound on the two ears also affects the directional sensitivity, especially with low notes for which refraction is negligible.

Differences in sound intensity and in the time of arrival of the sound waves occur only when the source of sound is displaced horizontally out of the plane of symmetry containing the two ears, but not when it is displaced in a direction perpendicular to this plane. The latter displacement is hence much less perceptible, a point which must be borne in mind when deciding on where to place the loudspeakers. The latter should as far as possible be placed above the speaker and not to his side. Also, the distances between the individual projectors and between them and the speaker should not be too great, as otherwise they will be heard independently of each other instead of simultaneously. Their distances from the audience should not differ by more than about 60 ft., for no time difference in the arrival of sound should exceed about ${}^1/_{17}$ sec.

In the majority of auditoria the sound intensity is not entirely independent of the positions of the loudspeakers, for the audience at the back of long halls will usually receive a greater volume of sound the higher the loudspeakers are placed. (A sound wave which travels directly above the heads of an audience is considerably enfeebled by acoustic absorption). A compromise must therefore be made here between the need for placing the loudspeakers as close as possible to the speaker and the need for obtaining satisfactory intelligibility in the rearmost rows of seats.

Sound intensity

Intelligibility is determined by the ratio between the sound intensity and the noise level or reverberation; this relationship has been shown diagrammatically in Philips techn. Rev., **3**, 139, 1938, fig. 2. If the sound intensity is 15 decibels higher than the noise level, the intelligibility for nonsense syllables is still 70 per cent and simple sentences can be clearly understood. In the absence of disturbing noises the sound intensity must not drop below about 30 phons, otherwise the intelligibility immediately drops considerably.

The relationship between the loudness level L in decibels above the threshold value, for which a value of 10^{-10} micro-watts per sq. cm. is assumed, and the output N of the source of sound is given by the expression:

$$N = \frac{V}{T}\, 10^{\frac{L - 73.8}{10}},$$

[1]) If $R(\varphi)$ is the potential induced at an angle of incidence, $d\Omega$ the solid angle between φ and $\varphi + d\varphi$, then the mean response apart from a factor of proportionality is:

$$\frac{1}{4\pi}\int R^2(\varphi)\, d\Omega = {}^1/_2 \int_0^{\pi} R^2(\varphi) \sin\varphi\, d\varphi.$$

For the pressure microphone, the pressure-gradient microphone and the combined microphone $R(\varphi) = 1$, or $\cos\varphi$, or ${}^1/_2\,(1 + \cos\varphi)$ respectively.

where V is the volume of the auditorium in cub. m.,
T is the time of reverberation in sec.,
W is the output of the source of sound in microwatts,
L is the loudness level above 10^{-10} microwatts per sq. cm. This formula can be readily derived from the relationship between the output of a sound source and the energy density given in an earlier article (A. Th. van Urk, Philips techn. Rev., **3**, 65, 1938, *cf.* p. 70). A nomogram is given in *fig. 2* for the relationship between N and L, which indicates that with the normal sound intensity of the human voice ($N = 10$ microwatts) a loudness level of 45 phons (above the threshold of audibility) can be obtained in an auditorium having a volume of 10000 cub. m and a reverberation time of 1.5 sec. This is just sufficient to understand every syllable the speaker utters, provided the interference level does not exceed 30 phons. With a higher noise level electrical amplifying equipment must be employed.

The required output of the electrical amplifying equipment can be calculated by adding 15 decibels to the noise level of the auditorium which gives the sound-intensity level required; the output which the loudspeakers together with the speaker must furnish is then given by fig. 2. If the efficiency of the projectors is known the average power output of the amplifier can be deduced therefrom. The maximum undistorted output of the amplifier must be about 15 decibels higher [2]) than the mean value calculated in this way.

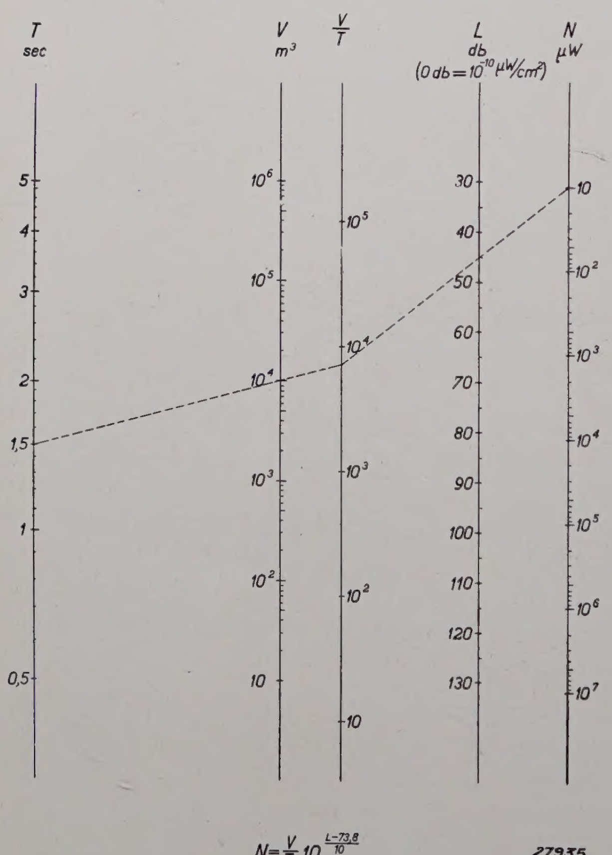

Fig. 2. Nomogram $N = \frac{V}{T}\, 10^{\frac{L - 73.8}{10}}$.

L = Sound intensity in decibels above 10^{-10} microwatt per sq. cm.
N = Output of sound source in microwatts.
T = Reverberation time in seconds.
V = Volume of auditorium in cub. m.

Sound amplification in buildings with defective acoustics

The production of echo and the focussing of sound waves in halls as a result of irrational planning and unsuitable dimensions have already been discussed in a previous article [3]). If the auditorium is too large so that an amplifying equipment must be used in any case owing to the inadequate sound intensity, these deficiencies in the design of the auditorium can to some extent be rectified by selecting suitable positions for the loudspeakers. In this way the amount of sound directed against walls responsible for undesirable reflection can be reduced to a minimum. Yet a really effective improvement in these cases is generally only possible by altering the plan of the auditorium or by covering the troublesome reflecting surfaces with sound-absorbent materials.

Another troublesome feature frequently found in churches, which is also due to the geometrical design of the auditorium and which can also be considerably remedied by means of an amplifying equipment, is the presence of many thick pillars in old churches which seriously obstruct the propagation of the high notes of direct sound waves. If the wavelength of the sound pulses is large compared to the dimensions of the pillars (lowest notes), the latter will not interfere with the transmission of sound. But if this ratio is small, as with high notes, the sound waves will be refracted, and in fact with very short waves a zone of silence may even be obtained behind the pillars. To illustrate how a zone of silence of this character is produced, the distribution of sound pressure about a spherical obstruction is shown in *fig. 3.* The ratio of the radius a to the wavelength λ is a measure of the refraction. The appearance of zones of silence of this character can be avoided by placing additional loudspeakers behind the pillars.

Another factor adversely affecting audibility, also frequently found in churches, is that the hearers

[2]) F. Trendelenberg, Klänge und Geräusche, J. Springer, Berlin, 1935. p. 92.

[3]) R. Vermeulen. Auditorium acoustics and intelligibility, Philips techn. Rev., **3**, 139, 1938.

are located both in front and to both sides of the preacher. The intensity of the sound waves is not the same in all directions owing to the refraction of sound waves round the speaker's head; the differences in intensity in different directions are greater for the higher frequencies than for the lower ones. Measurements of the directional distribution of the mean sound pressure of the human voice are reproduced in *fig. 4*, which shows that the intensity behind the speaker is about 14 decibels lower than straight ahead. Furthermore, the preacher frequently occupies a pulpit closed at the rear which intensifies the screening effect in this direction. An improvement in audibility can also be obtained here by placing a loudspeaker behind the pulpit.

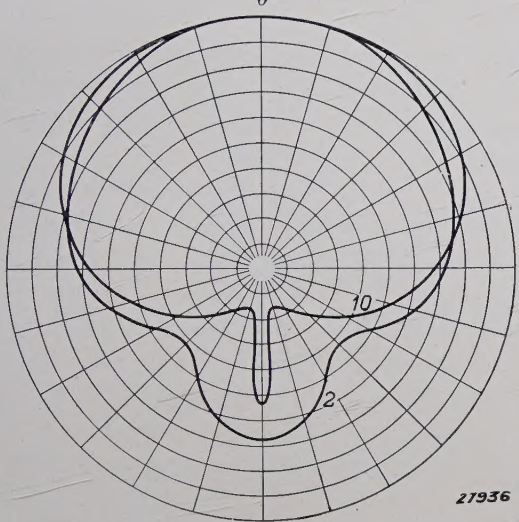

Fig. 3. Zone of silence behind an obstruction. Distribution of the sound pressure produced by a plane sound wave of wavelength λ at the surface of a sphere. The figures against the different curves are the values of the product ka, where $k = 2\pi/\lambda$ and a is the radius of the sphere.

The screening effect of the head can be calculated by replacing it by a sphere with a point source of sound located at its surface. Instead of calculating the sound pressure at a considerable distance in front of the sphere as a function of the direction, the pressure in a specific direction can be calculated as a function of the position of the source of sound on the surface of the sphere. But this relationship is identical with the distribution of the sound pressure at the surface of the sphere when the source of sound is far removed; this distribution of sound pressure has already been given in fig. 3. The reciprocal theorem employed here states in general terms that the sound pressure remains the same when the source of sound and the point at which the pressure is measured are interchanged.

Sound amplification in auditoriums with an excessive reverberation time

A discussion of the dependance of intelligibility on the ratio of the "useful" sound (*i.e.* the sound which reaches the hearers within an interval of $^1/_{17}$ of a sec.) to the reverberation has already been given in a previous article (Vermeulen, *loc. cit.*). If this ratio is too small the speaker's utterances become unintelligible. Knudsen has analysed the connection between intelligibility and

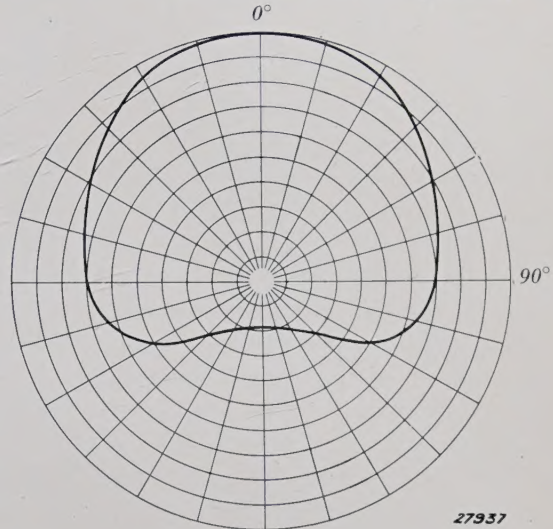

Fig. 4. Directional distribution of the mean sound intensity of the human voice according to O. Zwikker (Ingenieur, **44** E, 39, 1929). The curve shows the effective acoustic pressure in different directions. The acoustic pressure is about 14 decibels lower behind the speaker's head than directly in front.

the reverberation time, his results being reproduced in *fig. 5*. If the reverberation time exceeds 3 secs, the speakers's voice is as a rule unintelligible, and to obtain an improvement in intelligibility the acoustic pressure of the direct sound pulses must be increased but without a comparable increase in the reverberation. This may be realised by using loudspeakers with pronounced directional characteristics and which direct the total sound energy in the direction of the audience. The reflection of

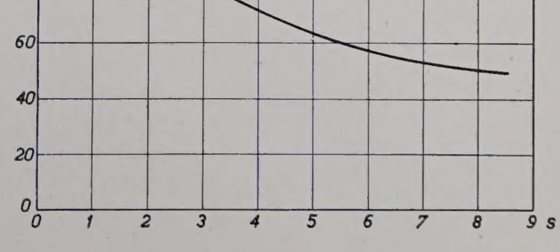

Fig. 5. Intelligibility as a function of the reverberation time according to Knudsen.

sound at a surface occupied by an audience is almost zero, so that the reflected sound waves are not amplified and hence reverberation does not become more marked.

Directional characteristics of loudspeakers

Whether the sound waves are propagated in a specific direction or not depends on the ratio of the dimensions of the radiating surface to the wavelength of the radiated sound pulses, as well as on the shape of the diaphragm and its method of mounting. If the wavelength is large compared to the dimensions of the radiating surface, a spherical sound wave is emitted which has the same intensity in all directions. On reducing the wavelength and making it more of the order of magnitude of the dimensions of the diaphragm, a pronounced directive effect may be obtained; the sound is radiated principally in a direction perpendicular to the radiating surface, interference between sound waves radiated from different points of the diaphragm taking place in other directions.

The distribution of sound can be accurately calculated for a plane, circular radiating surface which moves as a whole and which is mounted in an aperture in an infinitely large baffle, being given by the expression:

$$\frac{p_a}{p_0} = 2 \frac{J_1\ (ka \sin a)}{ka \sin a} \quad \dots \dots (1)$$

where $k = 2\pi/\lambda$; λ = wave length of the sound waves;

a = radius of the diaphragm;

p_a = sound pressure in a direction making an angle a with the normal to the surface of the diaphragm;

p_0 = sound pressure in the direction of the normal;

J_1 = first order Bessel function.

The directional distribution expressed by equation (1) has been plotted in *fig. 6* for various values of the parameter *ka*. This diagram shows that a satisfactory directional effect for a pulse of *e.g.* 500 c/s (k = 0.1 cm^{-1}) can only be obtained when a diaphragm 100 cm in diameter (ka = 5) is used. As a rule the diameter of a loudspeaker cone is only of the order of 20 cm and it is not practicable to make the diameter any larger. A larger diaphragm, which must also be sufficiently rigid to vibrate as a whole, would be very heavy, and this would detract from the efficiency of the loudspeaker as a whole.

The directional characteristics of a loudspeaker can however be improved in still another way, *viz.*, by placing the cone in a horn, the diameter of the radiating surface of the horn being larger than that of the cone so that a more pronounced directive action is realised.

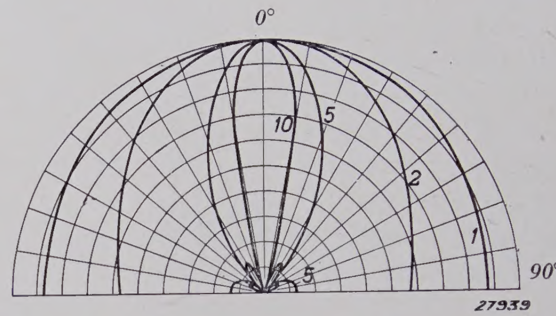

Fig. 6. Directional distribution of the sound intensity with a circular vibrating plate in an infinitely large baffle. The figures against the curves give the values of the parameter *ka* which determines the distribution.

If an attempt is made to use this method for diverting also the lowest audio-frequencies in a specific direction, the requisite dimensions of the horn would be so large that a practical design would prove entirely out of the question. If those frequencies which are not radiated in a specific direction by standard horns now in use are removed by an electrical filter in front of the loudspeaker input, satisfactory directional radiation can then be obtained for the whole sound output of the loudspeaker. The absence of the low frequencies will only slightly affect intelligibility (*cf.* Vermeulen, *loc. cit.*); if, for instance, all frequencies below 500 c/s are removed the intelligibility will still be of the order of 95 per cent. Yet, on the other hand, the filtering out of the low-pitched sound must not be taken too far, which would be obtained with hornless loudspeakers, as then the timbre would be altered too much and the audience become aware of a difference in timbre between the speaker and the sound radiated by the loudspeakers.

Fig. 7 shows the measured directional distribution of the radiated sound energy for a loudspeaker with a cone diameter of 22.5 cm, firstly with a baffle of 100 cm in diameter and, secondly, mounted in a horn with an aperture of 62 cm. It is seen that in the latter case the energy at frequencies above 500 c/s is radiated mainly within a cone of solid angle 60 deg. From the distance of the loudspaekers from the audience, the number of loudspeakers required for satisfactory diffusion of sound in a large auditorium can be immediately deduced. If the sound intensity level required is also taken into consideration, the acoustic energy which the loudspeakers must radiate can also be deduced.

The microphone should be placed as far as

possible outside the beam of waves radiated from the loudspeakers. The use of a pressure-gradient microphone is an advantage here and it should be so placed that the lines joining the microphone and the loudspeakers lie outside the response zone of the microphone.

Another method of intensifying the direct sound pulses in an auditorium relative to the reverberation consists in using a large number of small loudspeakers which are located at various points in the audience. Owing to the proximity between the loudspeakers and the audience the sound radiated by the projectors is entirely absorbed by the audience without any of the loudspeakers an auditorium. In many cases amplification of sound will therefore prove imperative. One way of doing this is by mounting a screen or sound reflector behind the speaker, although owing to the small dimensions of the reflectors used this arrangement is much less effective than reflection at the walls and particularly at the ceiling of an auditorium, and a really satisfactory improvement cannot be obtained by this means.

The intelligibility can be improved in this case by means of an amplifying equipment. If only a few large loudspeakers are to be used, their directional characteristics must be very good so that the available energy is utilised to the best advantage. When

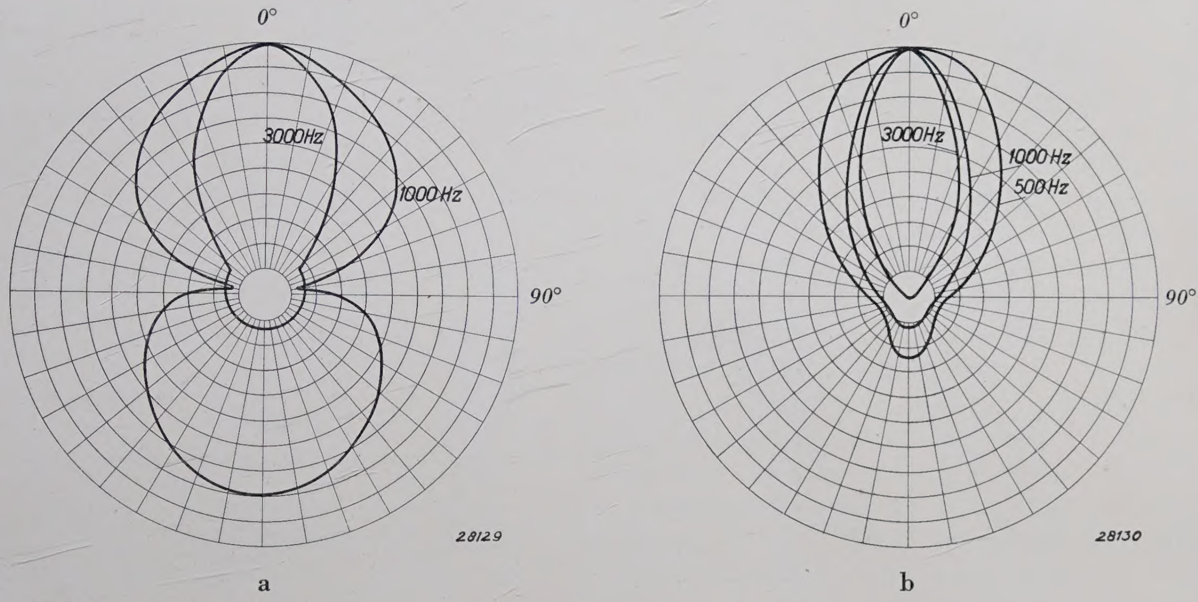

Fig. 7. Measured directional distribution of the sound pressure for a loudspeaker: a) when mounted in a baffle, and b) when mounted in a horn.

requiring to have a pronounced directional action. When using this method, care must be taken that the hearers do not receive any sound pulses from loudspeakers farther than 60 ft away, as otherwise the time difference in the arrival of sound pulses coming from the speaker direct and that radiated by the loudspeakers will begin to be disturbing.

Sound amplification in the open air

In the open air no reverberation likely to reduce the intelligibility will occur, although the level of other disturbing noises is usually higher than in an auditorium (sound from stands and public enclosures, the audience as a rule being far noisier than in an auditorium); the sound intensity required in the open air for satisfactory intelligibility is thus generally greater than that needed in deciding the direction of the loudspeakers, reflection of sound at the walls of the dais or stand should be carefully excluded.

The wind plays an important part in sound amplification in the open air. The wind intensity increases in a direction perpendicular to the ground, so that sound pulses which are propagated in the same direction as the wind are deflected towards the ground. But if the waves are propagated in the opposite direction. *i.e.* against the wind, they are deflected upwards away from the ground. In the first case, the range of the sound waves is greater and in the second case smaller than in still air.

With a wind of normal strength, the distance from the source of sound at which the spoken word is still intelligible may be double in the direction of the wind than in still air and only half as great when the waves travel against the wind.

AUTOMATIC MACROSCOPIC EXAMINATION OF MATERIALS WITH X-RAYS

by J. E. DE GRAAF and J. H. VAN DER TUUK.

539.26 : 537.562

A method is described for the automatic detection of flaws, *e.g.* cavities and blowholes, in castings, etc., using ionisation chambers and an amplifying arrangement.

Introduction; General arrangement of ionisation chambers

The disadvantage of the method of radiographic examination of macrostructure previously discussed in a series of articles in this review [1]) is that the cost per object examined is comparatively high; it is sometimes even greater than the value of the object itself in the case of cheap mass produced components, owing to the high cost of the films used and the length of time occupied for each examination. Instead of making a permanent radiograph, it is also possible to study the screened image visually on a fluorescent screen, this method of examination being frequently used for testing aircraft components, which usually have thin walls or are made of light metals, such as aluminium and magnesium. The method is convenient to apply and is quick, but has various disadvantages for it is not objective, does not give a permanent record, and it necessitates the observers being frequently relieved as the work is very fatiguing.

A method of examination is described in the present article which has been developed in our Laboratory and serves for the automatic indication of faults and flaws, such as cavities in castings. It should be emphasised at the outset that the apparatus described has not reached a stage of development fitting it for general adoption in industry, for its primary purpose up to the present has been the investigation of the fundamental possibilities of the method.

As a cavity in a metal object represents a centre of reduced absorption, the intensity of the X-ray beam which passes through a cavity will be greater than that of the same beam after traversing a solid path throughout, so that a measurement of the difference in intensity serves to indicate the presence and location of a cavity. The intensity of an X-ray beam can be measured with an ionisation chamber, the current flowing through this chamber as a result of ionisation being a measure of the beam intensity. But if only one ionisation chamber is used small fluctuations in mains voltage will also be indicated, since the radiation transmitted by a thick metal wall is proportional to a very high power (5 or 10) of the voltage applied to the X-ray tube, *i.e.* of the mains voltage. The effect of these fluctuations can be eliminated by using the compensation circuit shown in *fig. 1*. As long as the intensities of the two X-ray beams are equal the currents flowing through the two ionisation chambers a and b will also be equal, the current through the high resistance R will remain zero and the p.d. between c and d will also be zero. If the intensities of the beams are unequal, as for instance when one beam passes through a flaw in an object under examination, a p.d. is obtained between c and d. This p.d. can be employed for actuating an alarm arrangement connected to terminals e and f.

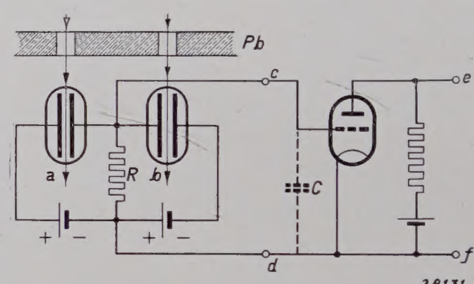

Fig. 1. Arrangement of two ionisation chambers a and b in a compensation circuit, enabling measurements to be made independent of simultaneous and equal fluctuations in the intensity of the two X-ray beams (fluctuations in mains voltage). A lead screen with two apertures for the passage of the beams is placed in front of the chambers.

The procedure of a test is as follows: The object under examination is moved between the ionisation chambers and the X-ray tube. In the case of a tubular object the best method is to rotate it about its axis and at the same time displace it along its axis. The X-ray beam thus scans the object along a helical path and any flaws present pass in front of the ionisation chamber at speed equal to the speed of revolution of the object.

This method can of course only be employed for detecting the presence of a flaw, a radiographic picture still being necessary to diagnose the cause, since only a photograph can give complete information regarding the shape, position and interrelationship of the flaws.

Sensitivity of the method

The difference in the currents through the two ionisation chambers and hence the p.d. between c and d (fig. 1) produced by a flaw in front of one of the ionisation chambers can be conveniently

1) Philips techn. Rev. **2**, 314, 350, 377, 1937; **3**, 92, 186, 1938.

calculated approximately when the X-rays are monochromatic. With an initial intensity I_0 and a thickness of material d or $d—\Delta d$ the intensities behind the object examined are (*fig. 2*):

$$I_1 = I_0\, e^{-\mu d},\ \text{resp.}\ I_2 = I_0\, e^{-\mu(d-\Delta d)},$$

where μ is the coefficient of X-ray absorption of the material. The ionisation currents produced in the ionisation chambers of length D by these X-ray intensities are:

$$i_1 = p\, I_1\,(1 - e^{-\mu' D})\ \text{resp.}\ i_2 = p\, I_2\,(1 - e^{-\mu' D}),$$

where μ' is the coefficient of absorption for X-rays of the gas in the ionisation chambers. The factor of proportionality p is partly determined by the wave-length λ of the X-rays, since with decreasing wave-length a diminishing portion of the energy primarily absorbed for the generation of a photo-electron is utilised for generating secondary electrons (ionisation). The ionisation chambers employed are in fact usually smaller than the range of the photo-electrons, produced by the absorption of the X-ray quanta, in the gas filling of the tube; these electrons thus reach the wall of the chamber before the whole of their energy has been consumed for ionisation and the unused fraction will be greater the greater the speed of the photo-electrons, *i.e.* the shorter the wavelength of the absorbed X-rays.

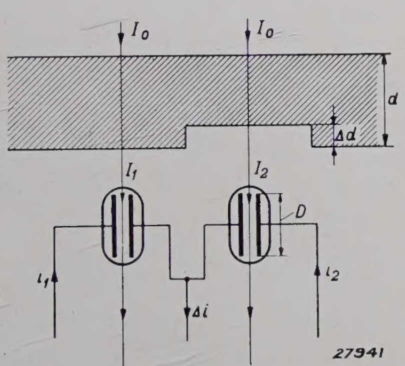

Fig. 2. Illustration of the calculation of the differential current. Δi due to a difference in thickness Δd in the object under test.

We thus get for the differential current:

$$\Delta i = i_2 - i_1 = i_1\left(\frac{i_2}{i_1} - 1\right) =$$

$$= i_1\left(\frac{I_2}{I_1} - 1\right) = i_1\,(e^{\mu \Delta d} - 1).$$

As $\mu \Delta d \ll 1$ (since Δd is very small), we have:

$$\Delta i = i_1\, \mu\, \Delta d \quad \cdots\cdots\cdots \quad (1)$$

The absorption in the thin-walled glass of the ionisation chambers and hence also the ionisation due to the photo-electrons emitted by the glass have been neglected here, as well as the effect of the radiation scattered by the object under examination; this is permissible as the X-ray beam used had a very small diameter.

The experimental relationship found between the strength of the signal S and the depth Δd of the flaw is shown in *fig. 3* for a test object with a wall-

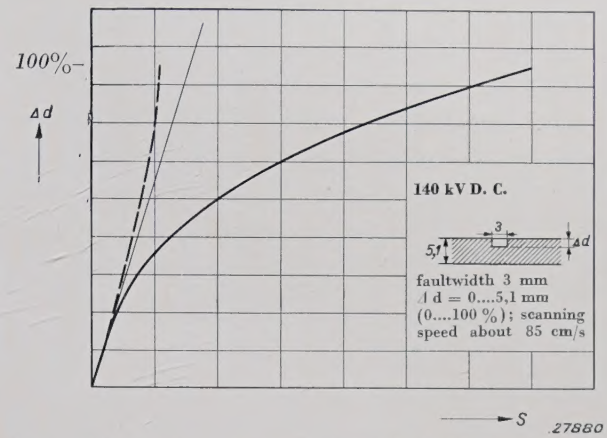

Fig. 3. Relationship between the depth Δd of the flaw expressed as a percentage of the wall thickness tested, and the signal strength S, in arbitrary units, with a flaw of constant width and a constant scanning speed.

thickness of 5.1 mm, the direct beam from an X-ray tube being used. According to equations (1) and their derivation, the curve would be expected to consist of an initial straight portion followed by a slower increase in the signal strength S with the depth of the flaw Δd, as for instance shown by the broken line in fig. 3. But in actual fact, S is found to increase rapidly with Δd, the increase commencing at a depth of about 30 per cent and approximating to the heavy line. This is probably due to the pronounced softening of the radiation transmitted by the flaw as Δd increases; towards the long waves this effect is less marked than with the rays transmitted through the overall wall thickness. The softer the X-rays, the greater will be their absorption by the gas filling in the tube and hence also the sensitivity.

There is a minimum current difference Δi which can still just be detected. At high amplifications the detection of flaws is limited by the statistical processes in the ionisation chambers, signifying that $\Delta i_{\min}$ is proportional to $\sqrt{i}$, *e.g.* $\Delta i_{\min} = q\sqrt{i}$. The smallest difference in thickness which can still be just detected in given circumstances is therefore:

$$\Delta d_{\min} = \frac{\Delta i_{\min}}{i_1\, \mu} = \frac{q}{\mu \sqrt{i_1}} \quad \cdots\cdots \quad (2)$$

The smallest relative difference in thickness is then:

$$\left(\frac{\Delta d}{d}\right)_{\min} = \frac{q}{\mu\, d \sqrt{i_1}} = \frac{q}{2\sqrt{p\, I_0\, \mu'\, D}}\, \frac{e^{\mu d/2}}{\mu d/2} \cdot \quad (3)$$

The quality of the X-rays (wavelength λ) is important in determining the smallest relative flaws which can be detected. If this relationship is investigated in the first place with a constant primary dosage ($p\, I_0 z \mu'\, D$ = const.) and variable wave-length, *i.e.* variable absorption coefficient, a minimum at $\mu\, d/2 = 1$ is found, in other words when about 12 per cent ($= e^{-2}$) of the initial intensity is transmitted. This optimum value is situated at a longer wave-length than that at which the thickness under examination is equal to the half-value thickness $d_{1/2}$, *i.e.* the thickness of the material at which 50 per cent of the initial intensity is transmitted. Since for this we have:

$$e^{-\mu d_{1/2}} = {}^1/_2, \text{ or } \mu = \frac{\ln 2}{d_{1/2}} = \frac{0{,}7}{d_{1/2}}\,.$$

it can therefore be concluded that with a constant wavelength (and constant primary dosage) the optimum value is obtained at a thickness which is several times greater than the half-value thickness for the radiation in question.

With a non-homogeneous beam of X-rays, equations (2) and (3) naturally apply only for each wave-length individually (with the corresponding values of μ and μ'), and in addition the primary dosage at constant tube current increases roughly with the square of the tube voltage. As a result of this increase the optimum tube voltage is displaced at constant thickness of material towards higher values than that at which 12 per cent is transmitted. Conversely, with a constant tension the optimum thickness approximates more and more closely to the half-value thickness, although always remaining above it. This indicates, *inter alia*, that the highest practicable tube tensions should be employed, because as most objects tested are 10 mm or more thick the half-value thickness after moderate filtering is only 6 mm. of steel when using a beam obtained with a load of about 400 kilovolts on the tube. The optimum value can thus be obtained only with the thinnest and lightest objects, and in any case higher tube voltages must be used than in radiography, where about 100 kilovolts are required for screening through 10 mm of steel *e.g.* in 3 minutes when using intensifying screens. The X-ray energy transmitted is then about 1 per cent of the primary dosage, instead of something over 12 per cent ($\mu d \sim 5$ instead of ~ 1).

The curve in *fig. 4* was obtained in an investigation of the validity of equations (2) and (3), in which the relative flaw detectible is seen to have a minimum in terms of the wall thickness d, when the ratio between the strength of the signal and the average of the statistical variations is always

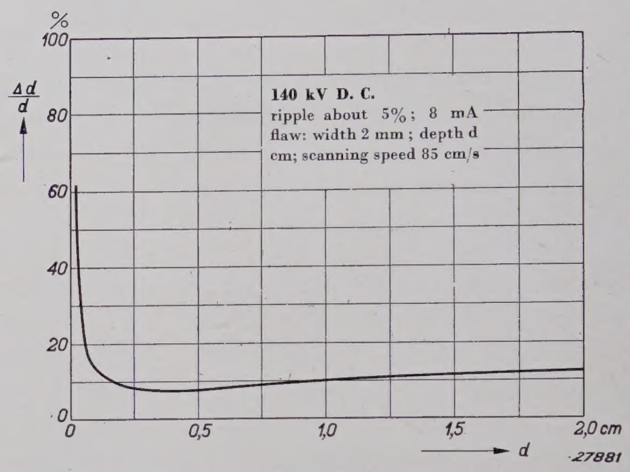

Fig. 4. The relative flaw $\Delta d/d$ detectible as a function of the wall thickness d at a constant tube voltage (140 kilovolts constant direct voltage; 8 milliamps), constant width of flaw (0.2 cm) and constant scanning speed (85 cm per sec.).

taken as equal to 2. The thickness $d = 0.4$ cm at the minimum, is, as can be expected, somewhat greater than the half-value thickness at the particular tube voltage applied (approximately 0.3 cm steel at 140 kilo-volts with the rays filtered as in the present case). The smallest detectible relative flaw is thus $7^1/_2$ per cent, in which connection it should be noted that, firstly, no steps were taken here to keep the statistical variations small, and, secondly, that a value of about 2 for the ratio of the signal required to the average statistical variation is on the small side for technical purposes. Double this value would be more acceptable.

Construction of the ionisation chambers

The construction of the ionisation chambers is shown in *fig. 5*, their principal features being exceptionally good insulation of the electrodes and rigidity, the latter being essential to prevent mutual vibration between the various parts of the

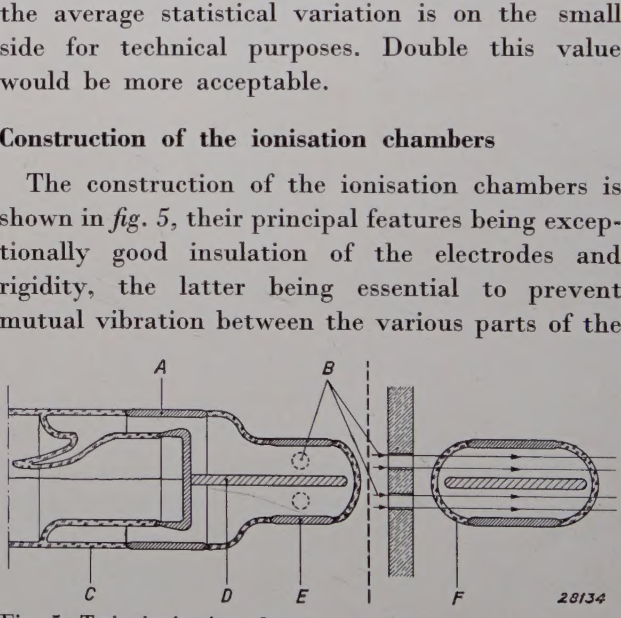

Fig. 5. Twin ionisation chambers for the automatic examinations of materials. A - earthed ring; B - beam of X-rays; the wall C is of a type of glass with very high insulating properties; D - exceptionally well insulated central electrode; E - electrodes connected directly to batteries. The wall F is made of a thin, very permeable type of glass.

apparatus (microphonic effect). The requisite insulation was realised by using a glass with a very low conductivity and by mounting an earthed ring A between the electrodes. All electrodes were made of chromium-steel so that they could be sealed through the glass. The middle electrode is a close fit in the ionisation chamber, but does not touch the glass so that the insulation is not in any way effected. The gas pressure in the two ionisation chambers is thus the same; they are naturally also of the same length.

The absorption of the radiation by the gas, and hence the magnitude of the ionisation current, increases with the atomic number of the gas and its pressure, xenon at a pressure of about 1 atmos. was used in the majority of the experiments, as well as in those to which fig. 4 applies. The pressure in a xenon ionisation chamber can be raised above 1 atmos. merely by allowing a known quantity of the pure gas to condense in the evacuated chamber, when the latter is immersed in liquid air. After sealing, the chamber is again raised to atmospheric temperature when the xenon evaporates. In this way a xenon pressure of 10 atmos. can readily be built up. The measurements described here with a load of 150 kilovolts applied to the tube showed that the number of ions produced in the xenon is roughly 50 times greater than in air at the same pressure.

The length of the ionisation chamber (D in fig. 2) should be made as large as possible, although it is shown below that no advantage accrues from increasing D with a specific cross-section if the capacity of the chamber is already large as compared with the residual capacity at the grid of the first amplifying valve. The diameter of the X-ray beam is made larger than the diameter a of the smallest blowhole to be detected, for these blowholes must give a just detectible signal, requiring that their whole area lie within the beam. Thus to examine a hollow cylindrical object, it is scanned by the beam along a helical path as already indicated above, the object being rotated about its axis and the chamber at the same time being moved parallel to this axis. If the pitch of the scanning helix is s, the optimum diameter of the beam will be $a + s$, since with a larger-diameter beam the signal given by the smallest flaw will not be any greater, although there will be an increase in the ionisation current in the rest condition and hence in the statistical fluctuations which limit the sensitivity. For the same reason the pitch should not be made too great and should be $s = {}^1/_3\, a$.

A diaphragm with two apertures should be placed between the object under test and the chamber to prevent scattered radiation from one beam reaching the other chamber, as this would reduce the differential current.

Statistical variations in the ionisation currents

The fact that the charged particles reach the electrodes at irregular intervals determines on the whole the magnitude of the minimum differential current detectible, since this current must be, for instance, three to five times greater than the average of the variations in the ionisation current. It is apparent that the statistical fluctuations of an ionisation current are considerably greater than those of an equivalent electron current emitted from an incandescent filament, since the fluctuations in current depend on the magnitude of the elementary charge. In the case of a radiating filament this is an electron, while in an ionisation chamber a very rapid primary electron, which generates a very large number (about 1000) of secondary electrons on collision with the gas atoms, is produced per absorbed X-ray quantum (in the intervals of statistical variation). On the average these collisions take place at equal intervals and in this way the whole yield of a primary electron more or less plays the part of the elementary charge in the statistical analysis of the current fluctuations. These fluctuations may in fact be 30 times ($\sim \sqrt{1000}$) greater than with a current of similar intensity emitted from an incandescent filament.

The average value of the fluctuations is proportional to the square root of the ionisation current [2]), so that as the ionisation current increases the smallest detectible flaw diminishes, when by taking the necessary measures for obtaining a high ionisation current (enlarging the chamber or raising the pressure) the differential pressure also increases proportionally (equation (3)). For a specific flaw the ratio of the signal to the fluctuations is then improved in proportion to the square root of the gas pressure in the chamber; increasing the current through the X-ray tubes gives the same result.

Differences in the velocities of the ions

The velocity of the carriers of the positive charges in the gas contained in the chambers, the xenon ions, is about 10 to 100 times smaller than that of the carriers of the negative charges in this gas. The gas used is in fact so pure that the electrons liberated cannot form negative ions with a much lower velocity by linking up with electro-negative

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[2]) Philips techn. Rev. **2**, 136, 329, 1937; **3**, 189, 1938.

atoms. The mean time of travel required by the positive ions to reach an electrode is about 3.10^{-3} sec., while that of the electrons is negligibly small. On a sudden change in the intensity of the X-ray beam, the associated change in the electron current in one of the chambers will occur about 3.10^{-3} sec. earlier at the middle electrode than the change in the ionic current in the other chamber. As a result a ripple in the tension applied to the X-ray tube, owing for instance to disturbances at the rectifier or insufficient smoothing of the tube voltage, will still give a signal in spite of using the compensation circuit in fig. 1. (Naturally this signal is much weaker than would be obtained from a single ionisation chamber with the ripple in question).

The high-tension unit

For two reasons the high-tension unit must furnish as constant a voltage as possible:

a) Because the detectible flaw depends on the voltage (I_0 and μ in equation (3)), and
b) because the ripple produces a signal.

A circuit must therefore be used which gives a constant direct voltage and which is provided with sufficiently large smoothing condensers. It is known from dosage measurements that the difference in the intensities through 10 and 7 mm. of steel increases with the fourth power of the tube voltage, when the latter lies between 100 and 200 kilovolts. An alteration in voltage of 5 per cent will thus result in an alteration in signal strength of about 20 per cent, indicating that the ripple in the tube tension must not exceed 1 to 2 per cent. A highly pulsating voltage is in any case quite useless, because time intervals then occur during which a flaw cannot be detected at all as no X-rays are passed through the object under examination.

Detectible flaws

According to equations (2) and (3) the detectible flaws are also determined by the thickness of the wall through which the rays are passed: the ratio of the relative flaws which can still be detected in 10 and 8 mm of steel is equal to 1.3 : 1, so that the ratio of the absolute flaws is 1.6 : 1. It was actually found in one case that flaws of 3 and 2 mm in depth could be located with equal facility in steel 10 and 8 mm thick respectively, which applies for a constant adjustment of Δi_{min}. The alteration in sensitivity with the wall thickness can be easily avoided by moving a template, which is also placed in the path of the rays, in unison with the object under examination. If the template has a suitable range of thicknesses the total thickness of material can be kept constant. A further advantage of this arrangement is that the ionisation currents, and hence also their fluctuations and the sensitivity of the apparatus (Δi_{min}), are kept constant irrespective of the thickness of the material. With a constant primary dosage ($p\, I_0\, \mu\, D$ = constant) the relative flaw would remain constant with a variable d, if the wavelength of the radiation is varied in such a way that the product of the coefficient of absorption and the thickness of the material d remains constant, in other words d is proportional to $1/\mu$. As for the half-value thickness $d_{1/2} = \log_e 2/\mu$ the quotient of thickness of the material and the half-value thickness $d/d_{1/2}$ would also be constant with a constant relative flaw. The curve of the half-wave thicknesses plotted as a function of wavelength, when drawn on a certain scale, will also represent the wall thickness with a constant value of the smallest detectible relative flaw as a function of the wavelength (with constant primary dosage).

First amplifying stage; form of signal

If a flaw, such as a blowhole in a casting, moves past one of the ionisation chambers a differential current with a time function as shown in *fig. 6* will

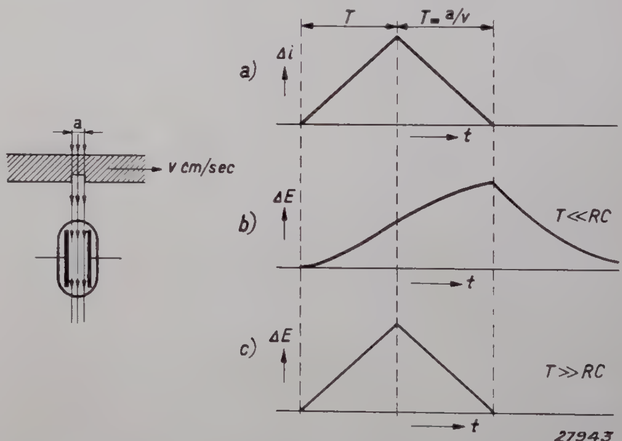

Fig. 6. Variation of current $\Delta i(a)$ through the grid resistance and the voltage ΔE (b and c) at the grid when a flaw of size a travels past one of the ionisation chambers with a velocity v.

be obtained, provided the diameter of the cavity is the same as that of the beam. Two limiting cases occur according to the circuit used, *viz.*, that the time T by the flaw to pass in front of the chamber is either very short or very long as compared with the relaxation period RC, where R is the grid resistance in the compensation circuit and C the total capacity at the grid (fig. 1). The variation of the grid voltage of the first valve in these two limiting cases is shown in fig. 6b and 6c respectively. The maximum resulting variations in grid voltage can be easily calculated. For $T \ll RC$, we have:

$$\Delta E_{\max} = \frac{1}{C}\int_0^T \Delta i\, dt = \frac{1}{C}\, T\, \Delta i_{\max} \quad \cdot \quad (4)$$

and for $T \gg RC$

$$\Delta E_{\max} = \Delta i_{\max}\, R \cdot \cdot \cdot \cdot \cdot \cdot \quad (5)$$

It may therefore be expected that at high scanning speeds v, to which equation (4) applies, the maximum strength of signal will be proportional to T (and inversely proportional to v). With very low scanning speeds the signal strength may be expected to be independent of the velocity v in accordance with equation (5). This is confirmed by *fig.* 7 for a particular case where $\Delta i_{\max} \sim 10^{-10}$

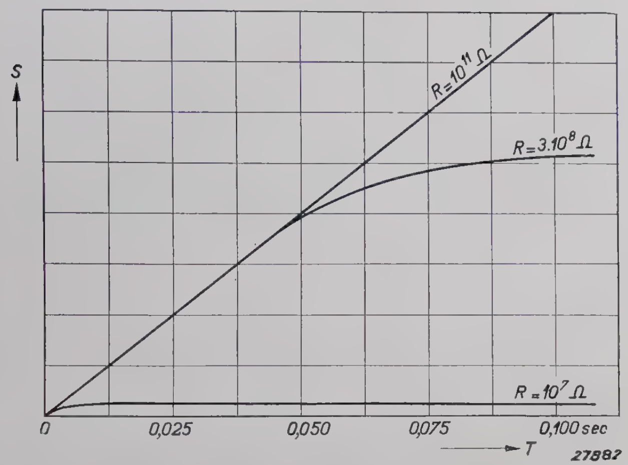

Fig. 7. Variation of the maximum signal strength S in arbitrary units with variable scanning speed v. The time T which the flaw takes to pass the chamber is plotted along the abscissa (see fig. 6); T is the quotient of the diameter of the flaw a and the scanning speed v.

amps. and $C \sim 10^{-10}$ farads. The greater the resistance R, the longer will the condition $T > RC$ be satisfied, and the longer will the curve remain a straight line and hence the higher will be the signal strength at which the curve bends to become horizontal.

For given values of $\Delta i_{\max}$, C and T the signal will become stronger with increasing R, until the condition in equation (4) is reached. A further increase in R is then useless, and in fact undesirable, as with a higher resistance there is a greater probability of interference occurring. The smaller C is made, the greater will naturally be the alteration in grid voltage. On making the ionisation chamber longer, $i_{\max}$ will increase in proportion to the length D (fig. 2), although at the same time the capacity of the chamber will increase, again in proportion to D. Under the conditions represented in equation (4) increasing the length of the chamber is useless, for as soon as the specific capacity of the ionisation chambers is made large as compared with the remaining grid capacity, then both Δi and C increase in proportion to D, and ΔE remains constant. In the case represented by equation (5) an increase in length will always give an improvement, although in general it will not be advisable to work with the conditions in this equation, since then ΔE is much smaller than with equation (4). An attempt could be made in the latter case to deduce the shape of the flaw from the shape of the ΔE curve, but an amplifier which will not distort arbitrary non-periodic voltage impulses is then required. In many cases, tests according to equation (5) offer an advantage, for instance in the examination of objects in which the various parts pass in front of the chambers at different velocities, *e.g.* tubes with different diameters which are rotated with a constant angular velocity; in these cases the smallest detectible flaw must not be determined by the velocity of motion of the object as required by equation (5). A better method in these cases is to adapt the speed of revolution of the tube to the diameter of the section passing in front of the chambers at each particular moment. It should also be pointed out that the maximum variation in voltage when $T \ll RC$ is a measure for the volume V of the cavity in the path of the rays, since:

$$\Delta E_{\max} = \frac{1}{C}\int_0^{2T} \Delta i \cdot dt\,(:)\int_0^{2T} \Delta d \cdot dt\,(:)\int_0^{2a} \Delta d\,\frac{ds}{v}\,(:)\int_0^{V} \frac{dV}{v},$$

where s is the distance traversed by the flaw. A necessary condition for this relationship to obtain is that the flaw is displaced at a constant speed v. When $T \gg RC$ the maximum variation in voltage is a measure for the maximum size of the flaw in the direction of the rays. These characteristics of the limiting cases (4) and (5) will also determine which conditions should be selected according to the circumstances ruling. These rules also apply approximately after further amplification of the signal.

When the flaw passes the two chambers in succession where both lie along the line of the scanning motion, a signal will be obtained first in one sense, *e.g.* an increase in grid voltage, and then in the opposite sense. The first signal must actuate the alarm device, which may be achieved by a reduction in a given voltage. To detect cavities in an object only one direction of motion is correct with a given amplifier, while to detect increases in thickness the opposite direction of motion is usually required. Difficulties can naturally arise with this arrangement, when the flaw is located in front of one of the chambers for so short a time that the times of transit of the ions are no longer

small compared with the rate of displacement of the flaw itself.

An electrometer triode, Type No. 4060, was used as the input valve, since this type of valve allows a very much greater grid resistance than other types. The voltage amplification was approximately 1 : 1 and the current amplification about 10^6. To protect the valve adequately against mechanical and electrical disturbances, it was enclosed in a heavy block of iron, itself mounted on resilient rubber feet and surrounded by a copper enclosure with walls 2 mm thick. The sensitivity of the valve itself is limited by the current fluctuations in the grid circuit in the absence of ionisation current. Alfvèn [3]) has calculated and measured a maximum charging sensitivity of approximately $2.5 \cdot 10^{-17}$ coulomb for an electrometer triode of this type. Measurements of the variations and an estimation of the grid capacity of the valve (approximately 10^{-11} farad) gave nearly the same value for the apparatus described here. Yet these variations are not a measure of the sensitivity for the present combination of ionisation chambers and electrometer valve, for the sensitivity is on the other hand determined by the variations in the ionisation currents flowing in the rest condition.

Further amplification and alarm device

For detecting small blowholes in castings which during the test must be moved rapidly in front of the chambers, the apparatus must be capable of indicating short voltage impulses; this allowed an ordinary alternating-voltage amplifier with condenser-resistance coupling to be used as a further amplifying stage. A single-valued relationship was found between the strength of the signal at the output of the amplifier and the size of the blowhole. The amplification was about 10^4.

The alarm arrangement consisted of a relay valve which is normally blocked by an appropriately high negative grid bias. A voltage impulse which reduces this grid bias friggers the valve, so that the anode current then switches on a siren or a lamp through a relay. By a suitable choice of the negative grid bias of the relay valve adjustment can be made to the minimum size of flaw detectible. The valve does not then allow current to pass with flaws smaller than this limiting value and no alarm is given. The alarm only ceases when cut off by the attendant (interrupting the anode current), so that no flaw can pass undetected owing to inattention on the part of the operators.

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[3]) H. Alfvèn, Z. Physik, **99**, 24, 1936.

Temporary design of apparatus and potential improvements

A temporary apparatus was set up for detecting blowholes not less than 3 mm in diameter in cast pipes with walls 10 mm thick. The ionisation chambers were placed inside these pipes, which were then scanned along a helical path (*fig. 8*). A special

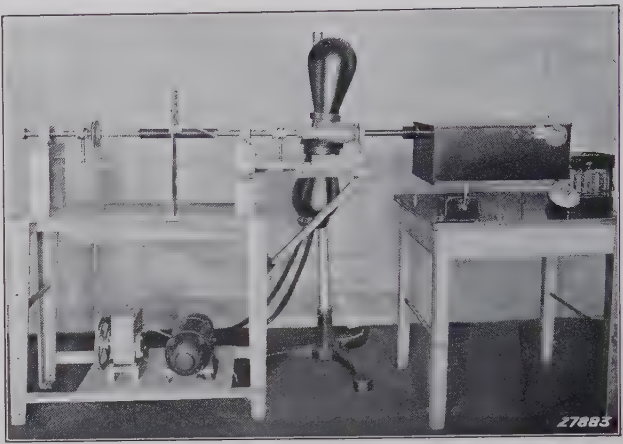

Fig. 8. The apparatus is adjusted for detecting blowholes 3 mm in diameter in cast-iron pipes with walls 10 mm thick.

machine was used for rotating the pipes and for moving them parallel to the axis. The speed of revolution was 200 r.p.m., and the pitch of the scanning motion 1 mm. The diameter of the scanning beam was 4 mm. A pipe length of 20 cm. was tested per min, the peripheral speed of the pipe being about 1 m. per sec. The walls of the pipes were 8 and 10 mm thick. If the apparatus is so adjusted that flaws of 3 mm are still just detectable in the 10 mm wall, an alarm is obtained with the 8 mm wall when the flaw measures only 2 mm. This undesirable alteration in sensitivity was corrected by a template of variable thickness which was moved together with the pipe, so that the total thickness of metal radiographed was maintained at 10 mm. Alternatively the negative grid bias of the relay tube could have been varied in direct relation to the position of the object holder. A variable diameter in the object under test can be allowed for by automatically altering the speed of revolution.

The apparatus is enclosed in a sheet-copper housing, which acts as a protection against external disturbances. The batteries for the ionisation chambers and the input amplifying valve are also accommodated in the same housing.

The characteristics of the apparatus described were adapted to the requirements specified, which were not very severe. Various features can still be considerably improved:

a) The ionisation chambers were about 15 cm from

the tube focus, but by using a modern small-diameter X-ray tube this distance can certainly be reduced to one half, thus increasing the intensity by a factor 4.

b) The current rating of the tube used can be at least doubled.

c) The gas pressure in the ionisation chamber can be increased to about 30 atmospheres, by means of which it can be safely assumed the ionisation current will be increased tenfold.

d) In very many cases the ionisation chambers can be made much longer, *e.g.* 60 mm instead of the 12 mm used above, *i.e.* five times greater. By using the X-ray tube referred to above, in which the focus is at the end of a narrow tube, the ionisation chambers can in most cases be placed outside the objects under examination, and the tube itself inserted within the object. The length of the chambers is then no longer determined by the internal diameter or bore of the object under test.

It thus appears possible to increase the intensity of the ionisation current about $4 \times 2 \times 10 \times 5 = 400$ times, allowing the thickness of wall of the object under test to be increased to possibly 30 mm. These improvements do not take into consideration any increase in the tube voltage, which as indicated above can be taken very high.

RADIO INTERFERENCE

by L. BLOK. 621.396.823

This article discusses how radio interference is produced by electrical apparatus and how the interference is impressed on a radio receiving set. The cause of radio interference is always a more or less discontinuous variation of voltages and currents in an electrical apparatus or appliance. Interference is transmitted to the receiving set mainly through capacitive coupling of the interference-carrying mains with the aerial or the aerial lead-in or by propagation of the interference through the mains to the mains unit of the receiver. In both cases the way in which the interference is transmitted from the producer to the mains is an important consideration and is discussed in detail. A distinction must be drawn between symmetrical and asymmetrical interference producers, the latter in particular being responsible for serious interference with radio reception.

Introduction

The increasingly severe demands imposed on the quality of radio reception have led to more and more attention being devoted to those factors which adversely affect and detract from the efficiency of reception of broadcast radio transmissions.

Distortion has been reduced to a minimum by progressive improvements in the construction of transmitters and receiving sets and it is, therefore, not surprising that as a result increasing attention has been given to the problem of interference with radio reception, whether due to atmospheric causes or to any kind of electrical machinery or appliance. This desire for purer reception has of course resulted in extraneous noises which were formerly accepted without complaint now being regarded as a serious disturbance. In consequence international efforts have been made to prohibit by legislation the use of electrical apparatus and appliances liable to cause interference with radio reception. The present article discusses the following aspects of this subject:

1) The production of interference by certain electrical apparatus:
2) The manner in which a disturbance is propagated from its source to a radio receiver; it is important in this connection to distinguish between symmetrical and asymmetrical components.
3) How a disturbance enters the receiving set.

Production of radio interference

Interference with radio reception is always due to more or less discontinuous variations in current and voltage. The clicking sound produced in a loudspeaker by the operation of a switch in a receiving set irrespective of the wavelength to which the receiver is tuned, is a well-known example of this. The cause of this noise is that a current impulse is composed of a frequency band in accordance with the Fourier integral theorem, which may contain all frequencies from zero to a value which is higher the more abrupt the impulse. As the receiver responds to all waves in the frequency band to which it is tuned, the clicking sound is produced.

If the current impulses occur periodically, as

those originating in electrical motors, rectifiers, electro-medical apparatus, etc. the disturbance will also be repeated periodically and will then become audible as the crackling sounds familiar to all listeners.

In many commutator motors disturbances are caused by voltages produced as the brush moves from one commutator bar to the next. If the brush remains in contact with two consecutive bars that part of the armature winding between these two bars is short-circuited. The voltage between the bars and the brush is then zero, but as soon as the brush breaks contact with one of the bars the short circuit over part of the armature winding is suddenly removed, so that the short-circuit current rapidly drops to zero. This action is accompanied by a rapid increase in the p.d. between the bar and the brush, which may give rise to more or less intensive sparking. When the brush comes in contact with the next bar another section of the armature winding is shorted and a short-circuit current is again rapidly built up, and so on. These sudden variations in current induce voltages over a wide frequency band in the whole of the armature winding, and these pulsating voltages may be impressed on the mains to which the motor is connected and in this way cause interference with radio apparatus fed from the same mains or situated close to a conductor in the mains network.

It should be noted that sparking is not essential for interference to be produced. If a motor does not spark, this does not justify the conclusion that the motor cannot interfere with radio reception.

Other pieces of apparatus which owing to poor design may interfere with radio reception are rectifiers. Abrupt changes in voltage occur in gas-filled rectifiers owing to the difference between the re-ignition voltage V_D required in each half-wave and the running voltage V_B. The greater this difference V_D-V_B, the more powerful will be the disturbances caused by the apparatus. Exceptionally high values of V_D-V_B may occur in controllable rectifiers with relay valves in which V_D is dependent on a variable grid voltage [1]).

Other electrical apparatus liable to act as sources of disturbance include discharge tubes used for advertisement signs and electro-medical apparatus, which generate high-frequency electrical fields.

[1]) The grid voltage varies with the same frequency as the voltage at the anode of the rectifying valve. The D.C. can be controlled by carrying the mutual phase displacement of these two voltages. *Cf.* the detailed discussion by J. W. G. Mulder and H. L. van der Horst: A controllable rectifying unit for 20 000 volts and 18 amps. Philips techn. Rev., 1, 161, 1936.

Propagation of interference

The disturbance generated can be impressed on radio receivers in various ways:

a) by direct radiation or by capacitive coupling of the source of disturbance itself with the aerial or the aerial lead-in of the receiver;
b) by radiation or by capacitive coupling of lighting mains propagating a disturbance, with the aerial or the aerial lead-in;
c) by transmission of the disturbance through the mains direct to the receiving set;
d) by a combination of the above.

A disturbance rarely reaches a radio set by the means enumerated under a), in fact only when the aerial is located close to the source of the disturbance. In practice radiation is as a rule quite small. Screening of the source of disturbance is the only method of suppression which can be adopted with this type of interference, if the cause of the disturbance itself cannot be eliminated directly at the source.

Far more frequently, and in fact nearly always, interference with radio reception is due to one or another of the methods enumerated under b) and c). To obtain a closer insight into the conditions ruling in these cases we shall investigate how a disturbing voltage can be impressed on the low-voltage mains and how it reaches a radio receiving set through these mains. An equivalent layout has been evolved below in which the circuits affected by the interfering voltage are indicated.

Equivalent circuit of a disturbance source

The interfering voltage of an electrical apparatus can in general be represented as a varying electromotive force E with an internal impedance Z_i in series with it (see *fig. 1*). If the apparatus is connected to the mains, both terminal *1* and terminal *2* of the interfering voltage are connected through certain impedances with the mains terminals *3* and *4*. In addition a certain impedance will exist between terminals *1* and *2* and between terminals *3* and *4*, so that a generalised quadripole (V_1) is produced (see component *I* in fig. 1). Frequently certain parts of the electrical apparatus are earthed or possess an appreciable capacity to earth. As a result terminals *1* and *2* are connected to earth through certain impedances Z_{1a} and Z_{2a}. All these impedances are of course, usually complex and dependent on the frequency.

Having determined which components of the circuit transmit the interfering voltage to the mains terminals *3* and *4*, we shall discuss the component marked *II* in fig. 1 which connects the mains

terminals of the interference source to the mains terminals 5 and 6 of the receiving set. If consideration is limited to terminals 3, 4, 5, 6 and to the earth A of the whole complex network — the importance of the earthed pole will appear later on in this discussion — the mains circuit together with the connected apparatus (earthed or not) can be replaced by a generalised quadripole N and four impedances Z_3 to Z_6, which link the four connections of the quadripole with earth [2]).

Finally, component *III* of fig. 1 gives an equivalent circuit for the receiver which is composed of an impedance between the terminals 5 and 6 and two earthing impedances Z_{5a} and Z_{6a} between these terminals and earth.

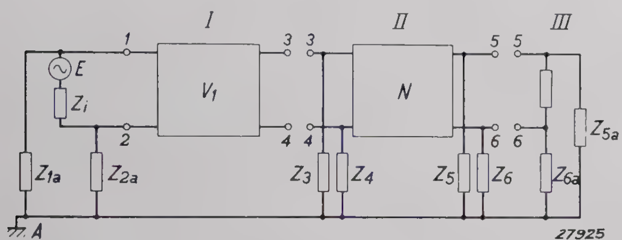

Fig. 1. Equivalent circuit of an interference source. The interfering voltage E with the internal impedance Z_i is located between terminals *1* and *2*. Terminals 5 and 6 are the inputs of the receiver; A is earth. V_1 and N are generalised quadripoles; the first connects the interfering voltage to the mains terminals of the interference producer, and the second represents the network between the interference source and the terminals of the receiver.

Symmetrical and asymmetrical interference

Before referring in detail to the various components of the circuit outlined above, our diagram must be further simplified in order to show that a disturbance can be transmitted in two fundamentally different ways from terminals *1* and *2* to terminals *5* and *6*. We here make use of the fact that an e.m.f. E with an internal impedance Z_i in series with it is electrically equivalent to a current source of intensity $I = E/Z_i$ and an impedance Z_i in parallel with it. This is shown diagrammatically in *figs. 2a* and *b*. The three impedances in fig. 2*a* form a triangle, which can be electrically replaced by a star circuit as shown in fig. 2*c*.

The simplification referred to consists essentially in substituting for the whole circuit between terminals *1*, *2* and *5*, *6* a single quadripole whose terminals are earthed through given impedances. These impedances are connected in parallel to the impedances Z_{1a}, Z_{2a}, Z_{5a}, Z_{6a} shown in fig. 1 and to both sides of the quadripole form a triangle of impedances between the points *1*, *2*, *A* and *5*, *6*, *A* respectively. On transforming this triangle into a star circuit according to fig. 2 and replacing, also as in fig. 2, the interfering voltage by an interfering current, a new equivalent circuit is obtained which is shown in *fig. 3*.

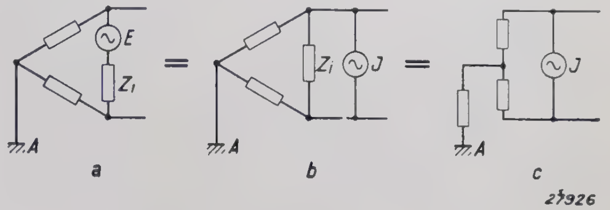

Fig. 2. Three equivalent circuits for an interference source. The voltage source can be replaced by a current source, and the impedance triangle can be transformed into a star circuit.

The interfering current I will flow partly through the impedances Z_1 and Z_2 and partly through the junction between terminals *1* and *2* in the mains circuit V (for instance, through the capacity between the mains conductors) and finally through impedances Z_5 and Z_6 in the receiver. In general the midpoints P_1 and P_2 of the star circuits will not be at the same voltage, so that part of the interfering current will flow through the earth connection. The impedance $Z_1 + Z_2$ can, however, always be subdivided into two impedances Z_1' and Z_2' (see *fig. 4a*), such that point P_0 is at the same voltage as the star point P_2 in fig. 3. A potential difference will exist between P_0 and P_1 which will be in a constant (complex) relationship to the interfering current I. We shall term this potential difference the asymmetrical interfering voltage and denote it by the symbol E_a. If we impress an e.m.f. equal to the asymmetrical interfering voltage $E_a = V_{P_1} - V_{P_2}$ as shown in fig. 4*b*, no p.d. will exist between the terminals P_0 and P_0' in this diagram. An opposing voltage source will be connected in series with this e.m.f., so that 4*b* really coincides with 4*a*. By the principle of superposition 4*b* can be resolved into two circuits as shown in fig. 4*c*. The voltages and currents at the receiver are equal to the sum of the voltages and currents

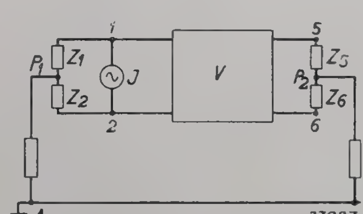

Fig. 3. Equivalent circuit of an interference source as in fig. 1, after simplification by introducing the transformations shown in fig. 2.

[2]) Regarding this point and other general characteristics of linear networks *cf.* the article by Bath. van der Pol and Th. J. Weyers, Electrical filters, Philips techn. Rev. 1, 240, 270, 298, 327, 363, 1936.

produced by the interference sources *I* and *II* in fig. 4*c*.

This resolution has the following advantage: It is seen that although impedances are present in component *II* of the interference source along the earth connection between P_0 and P_2, no potential differences are yet obtained. The current in the earth conductor is therefore zero. A voltage E generates a current which produces the same potential difference along both mains conductors (and thus has the same direction and roughly the same magnitude in both conductors), and which flows back through earth. The voltage between the mains conductors is usually subject to only a slight change as a result of this, the principal effect being voltage fluctuations between the

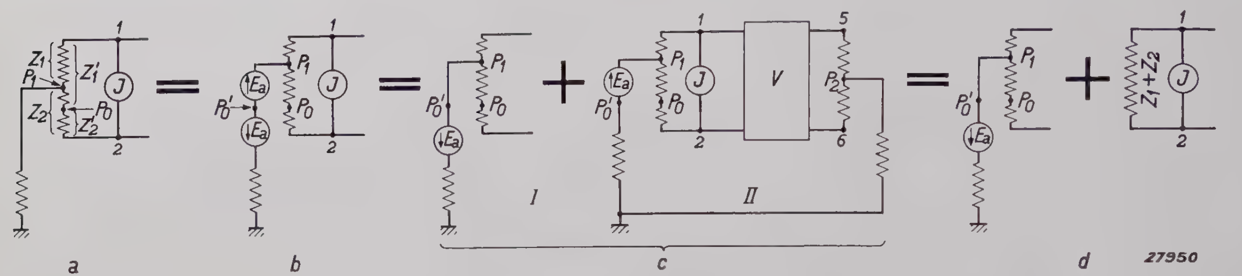

Fig. 4. Four similar equivalent circuits of the input in fig. 3. By introducing an asymmetrical e.m.f. E_a in such a way that the potential of P_0' is made equal to that of P_0, the source of interference can be subdivided into two components *I* and *II* (fig. *c*) of which the symmetrical component *II* produces no flow of current to earth.

part of the aggregate interference is thus isolated which produces currents of equal intensity but of opposite polarity in the two mains conductors, so that no currents flow to earth. This component we shall term the symmetrical source of interference. The earth connections need not be considered for a closer investigation of the effect of the symmetrical interference component, so that the equiualent circuit can be further simplified to give that shown in fig. 4*d*. Finally, by eliminating the earth connection the current source with the impedance $Z_1 + Z_2$ in parallel can be again replaced by a voltage source $E_s = I\ (Z_1 + Z_2)$ with the impedance $Z_1 + Z_2$ in series. The new equivalent circuit for the complete disturbance is shown in *fig. 5*.

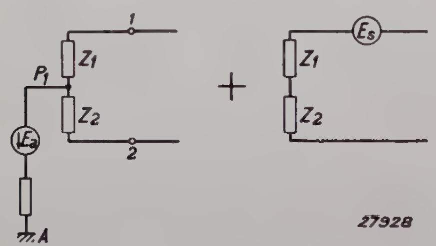

Fig. 5. Simplified layout of the symmetrical interference component E_s and the asymmetrical interference component E_a, which together furnish the interference voltage as shown in fig. 1.

The symmetrical interfering voltage E_s produces opposing currents in the two mains conductors, no potential differences appearing between the two star points and earth. The asymmetrical interfering voltage E generates ... conductors and earth. These two interfering voltages can now be usefully discussed in further detail.

Symmetrical interfering Voltage

The equivalent circuit for the symmetrical interfering voltage is shown in *fig. 6* in somewhat more detail than in the previous figure. Z_i is the internal impedance of E_s; the characteristics of the mains are for the sake of simplicity repressented by three impedances [3]). The magnitude of Z_1 (which can be both inductive or capacitive) is dependent on the frequency.

The mains impedances, Z_s and Z_p, are also functions of the frequency, being governed by the length and layout of the network. Measurements have shown that the two mains conductors usually have a very high capacity with respect to each other so that the impedance Z_p is small, and as a

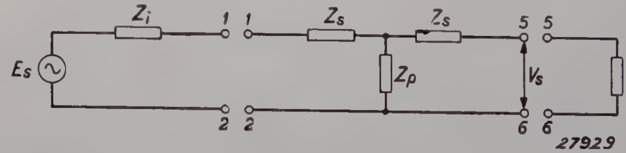

Fig. 6. A detailed scheme of the symmetrical interference component. The interfering voltage V_s is generally small compared with E_s, because the parallel impedance Z_p between the mains conductors is usually very small owing to their mutual capacity.

[3]) This is possible as the earth connections have been eliminated so that there is no external connection between terminals *1* and *2* and terminals *3* and *4*.
Cf. the article by Balth. van der Pol and Th. J. Weyers, *loc cit.*

result the interfering voltage V_s at the receiver is usually very small as compared with E_s, provided Z_i does not fall below a certain lower limit. A symmetrical interfering voltage in the mains will thus as a rule cause little interference in radio receivers connected to these mains.

As the two mains conductors are generally situated close together, the radiation from the mains is also small where a symmetrical interfering voltage exists.

Asymmetrical interfering voltage

As already indicated above, a circuit with an asymmetrical interfering voltage is made up of the two mains conductors connected in parallel and the earth connections, according to the arrangement shown in *fig. 7*.

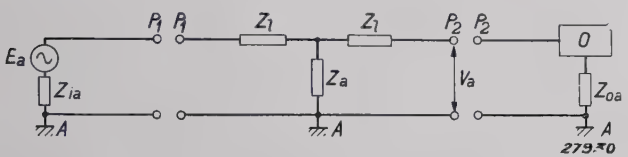

Fig. 7. Layout of the asymmetrical interference component. The interfering voltage V_a may be much greater then V_s (fig. 6), particularly when the earthing resistance Z_a of the mains is not too small.

The interfering voltage E_a is connected to earth through the impedance Z_{ia}. The series impedances of the mains conductors in parallel are represented by Z_l, and the capacitive impedance between mains and earth is denoted by Z_a.

In the case of cables the capacity of the cores to earth is usually fairly high; not only has a cable section buried in wet soil a high capacity to earth, but the unburied sections also exhibit a pronounced capacity between the cores and the lead sheath, the latter being more or less well earthed. Thus Z_a will usually be small. But in the conductors of house installations Z_a may have a much greater value. Although the tubing or lead sheath surrounding these conductors is frequently earthed at one point, this is not sufficient. Close to the earthed point the tubing or the lead sheath may be regarded as equivalent to earth; at some distance from this point the resistance of the tubing cannot be neglected, especially as it increases with the frequency as a result of skin effect. Owing to the considerable distance between the conductors in a house installation and earth, which in this case acts as a return, radiation from the conductors with an asymmetrical interference component is much greater than with a symmetrical disturbance, which accounts for the fact that in by far the majority of cases interference with radio reception can be ascribed to the asymmetrical components of the interfering voltage, when the source of interference is not situated in the immediate neighbourhood.

A feasible method for suppressing asymmetrical interference appears to be a modification of the receiver circuits so that the star point P_1 in fig. 3 is at the same potential as P_2. This is not altogether simple since the p.d. between P_2 and P_1 depends on the nature of the quadripole V between terminals *1* and *2* of the interfering source and terminals *5* and *6* of the receiver. But this quadripole contains all side branches of the network, and in these branches the nature or magnitude of the load may vary, which will result of course in a change of the characteristics of the quadripole V.

Yet if the earthing resistance Z_{ia} of the interfering source is made sufficiently large the asymmetrical current can be kept within specified limits. Z_{ia} will then be determined not only by the earthing resistance of the interference producer, but also by the earthing impedances Z_3 and Z_4 in fig. 1. This is apparent from the method by which fig. 3 has been derived from fig. 1. These impedances are usually small and as a result Z_{ia} is also reduced. The effect of the impedances Z_3 and Z_4 in fig. 1 on the earthing resistances Z_{ia} can however be reduced to negligible dimensions by making the series impedance of V_1 (fig. 1) high on the side of terminals *3* and *4*. In certain cases it is then possible to give Z_{ia} a high value.

Another method of suppressing interference consists in making Z_1 and Z_2 (fig. 3) very small, so that the interfering current I is shorted in the interference producer itself and no longer flows through the asymmetrical circuit.

Which of these methods of suppression will give the most practical solution in each particular case will be discussed in a subsequent article.

How interference reaches the Radio Receiver

We have still to investigate how interference is produced in a radio receiver itself. The principal cause is the capacitive coupling between the aerial or the aerial lead-in and mains carrying an interference characteristic. It may also be possible for the mains transformer, whose primary is connected to the disturbing mains, to transmit the interference to the input circuit of the high-frequency component.

The fundamental circuit of that part of the receiver responsible for interference is shown in *fig. 8*, with an equivalent circuit at the side. C_t is the

capacity between the primary and secondary windings of the mains transformer, Z_1 is the impedance of the tuned circuit, Z_{0a} the resistance of the earth connection, C_A the earthing capacity of the aerial and C_k the capacity between the aerial and the mains conductors.

The interfering voltage V_a is responsible for the amplifying valve as an interfering voltage.

If the capacity C_t between the primary and secondary windings of the transformer is not very small, the interfering voltage can also give rise to currents in two other circuits, *viz.*, those flowing:

3) through C_t, Z_a, A_1 to A_0; and
4) through C_t, Z_1, C_A, A_2 to A_0.

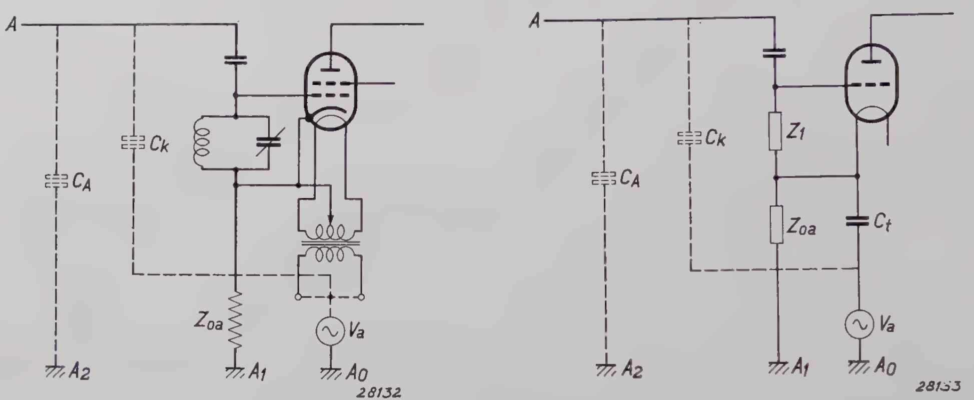

Fig. 8. Circuit of the input component of a receiver, with associated equivalent circuit. It is shown here how the asymmetrical interfering voltage V_a produces an alteration in voltage at the control grid of the high-frequency amplifying valve. This alteration in voltage is produced by currents which flow either through the capacity C_k between the aerial and the mains conductors or through the capacity C_t between the primary and secondary windings of the mains transformer.

appearance of a current in various circuits. Assuming that the capacity C_t is very small, as in modern receiving sets, then current will flow:

1) through C_t, C_A, A_2 to A_0; and
2) through C_t, Z_1, Z_{0a}, A_1 to A_0.

In the second case a voltage drop will appear at Z_1 and will reach the grid of the high-frequency The latter current will again produce an interfering voltage at the grid of the high-frequency amplifying valve. This form of interference would not occur if Z_{0a} were zero. The importance of a good earth for the receiver, in addition to a low value of C_t, in order to prevent interference reaching the set through the mains transformer is thus self-evident.

FLUORESCENCE AND PHOSPHORESCENCE

by J. H. GISOLF and W. DE GROOT.

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To supplement a previous article, the phenomena of fluorescence and phosphorescence (photoluminescence) of a series of different compounds are discussed here, *viz.*, lines and bands in fluorescence spectra of organic salts and organic compounds in solid and liquid solutions, and fluorescence and phosphorescence due to impurities, such as the luminescence of alkali-halide luminophors, ruby, sulphide luminophors, silicates, and fluorite. Tentative explanations regarding the mechanism of luminous emission in different cases are advanced on the basis of investigations of the afterglow.

Introduction

In a previous article [1]) the phenomenon of fluorescence was discussed with the aid of a few simple examples (resonance in monatomic and diatomic gases), some reference also being made to the fluorescence of solid substances (uranyl compounds) and liquids (eosin and fluorescein solutions), and to the so called luminophors. The difference between fluorescence and phosphorescence rests on the difference in the ability to re-emit selectively absorbed radiation, usually of a spectral composition different from that absorbed, in some cases only during the period of irradiation, *i.e.* fluorescence, and in other cases the persistence of an appreciable emission of light after irradiation has ceased, *i.e.* phosphorescence. Closer investigation has, however, shown that all bodies which can be rendered luminous by irradiation exhibit a finite afterglow period which may vary between a thousand-millionth of a second (10^{-9} sec.) and several months. These two phenomena are therefore frequently described as luminescence or more correctly as photo-luminescence to distinguish them from the emission of light under the action of cathode rays, X-rays, heat, mechanical deformation, chemical reactions, etc.

The phenomenon of photo-luminescence is quite common, in fact so widespread that with the exception of perfectly pure metal surfaces nearly all bodies can be assumed to luminesce when irradiated with light of suitable wave-length. A. van Wijk [2]) has already called attention to this in an article on the application of ultra-violet radiators for the investigation of luminescence. In lighting technology strongly luminescent bodies are naturally those to which the greatest interest attaches.

Luminescence may be a specific property of a certain atom or a certain atomic group or of a specific body in a pure and rarefied state, such as sodium vapour and iodine vapour discussed as examples in the previous paper. In many cases of this type the luminescence is retained in concentrated solutions or in the solid state; examples of this are salts of the rare earths, benzene, and uranyl salts. In other cases the luminescence of certain compounds is found only in the solid state, as in the platino-cyanides. But frequently, as seen in discussing fluorescence, a high concentration reacts adversely on luminescence, while dilute solutions exhibit a marked fluorescence, both in the liquid state (alcohol, water) and in the solid state (solutions in alcohol at —180 deg. C and in boric acid) and in the adsorbed state. Many bodies of organic origin probably owe their luminescence to that of impurities present in low concentrations. Inorganic luminophors whose luminescence is due to traces of certain admixtures, occur in Nature (ruby, fluorite) and can also be prepared artifically.

Although it is readily possible to co-ordinate the very diverse types of photo-luminescence in a rational classification, this does not signify that the mechanism of the phenomenon must be the same in all cases. On the other hand, a closer examination of the variation of the intensity of the afterglow with time has shown that *e.g.* the luminescence of a solution of rhoduline orange in sugar solution is due to an entirely different mechanism to that of a zinc-sulphide/silver luminophor. The various types of luminescence and the nature of the afterglow are discussed in further detail below.

Line and band fluorescence of crystals

Contrary to the behaviour of salts of other metals (alkalis, alkaline earths, copper, silver, etc.) the salts of the rare earths, both in the crystalline form as in solution, give small absorption bands in the visible and infra-red portions of the spectrum. In the solid state and at a low temperature (—180 deg. C.) these bands are seen to consist of a series of well-defined absorption lines with a definition comparable to that of the spectral lines obtained with gases at

[1]) Philips techn. Rev., **3**, 125, 1938.
[2]) Philips techn. Rev., **3**, 5, 1938.

low pressures. Irradiation with light having the wave-length of the absorption bands induces a fluorescence containing not only the same bands but also those with a longer wave-length. These absorption and emission bands have been classified in an energy-level scheme, an example of which is shown in *fig. 1*, *viz.*, for the trivalent terbium

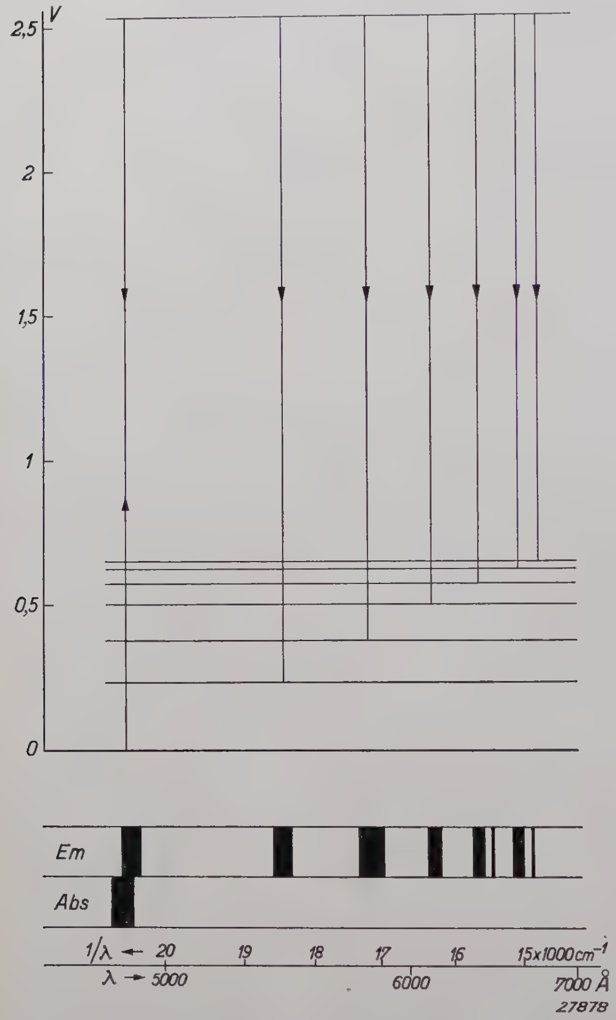

Fig. 1. Energy scheme of the ion Tb^{+++} and absorption and emission spectra of $Tb_2(SO_4)_3$, 8 H_2O according to Gobrecht (1937). Each of the transitions shown corresponds to a group of fine lines, each group being shown in the figure enclosed by a black rectangle. Particularly at low temperatures are these lines distinctly separated.

ion Tb^{+++}. The corresponding emission and absorption spectra are shown below, and the individual groups of lines are represented in the scheme as black rectangles. The resolution of each group into a number of fine lines (fine structure) is due to the action of electrical fields from the neighbouring ions in the crystal lattice. The most striking feature is that the width of each group is comparatively small (about 0.05 eV), *i.e.* much smaller than would be expected by analogy of the effect of electric fields on other atoms. Bohr's theory of the atom explains this phenomenon quite naturally, as in this theory corresponding energy-levels are ascribed to electronic orbits which lie at such a depth in the atom that they are only slightly affected by external electric fields, and which only in the case of the atoms of the alkaline earths occupy such positions that emission and absorption of light can be associated with them.

The uranyl salts also exhibit absorption and emission lines in solution and in the crystalline state, these lines being well defined in the solid state and at low temperatures. It has already been shown (Philips techn. Rev., **3**, 130 and 131, 1938) that here the oscillation of the oxygen atom with respect to the uranium atom in the radicle UO_2^{++} produces a band spectrum and that the important characteristics of this spectrum can be explained on simple lines.

Similar to the two cases just discussed, still a third type of band fluorescence can very probably be ascribed to a specific atomic grouping, *viz.*, that observed in the double cyanides of platinum with other metals. Photoluminescence has been observed in the platinocyanides of barium, magnesium, lithium, sodium, potassium and rubidium; thus the use of barium platinocyanide for screens in radiography is well known, although of course this application naturally does not rest on photoluminescence in the narrower sense. Similar to the absorption spectrum, the emission spectrum consists of very broad ill-defined bands occupying very different positions with the various compounds. This indicates that if the platinocyanide group ($Pt(CN)_6^{--}$) is the focus of luminescence, it is nevertheless strongly influenced by its surroundings.

It is extremely difficult to make quite certain whether an observed luminescence is due to the pure substance or to admixtures present in such very small quantities that they are not capable of analytical determination. In the case of the platinocyanides, it is not very probable, although perhaps not quite impossible, that the effect is due to impurities.

Other compounds are also known which luminesce only in the crystalline state and whose absorption and emission bands are regarded as most likely due to the pure substances; these include a number of tungstates and molybdates. Of these calcium tungstate, which on irradiation with ultra-violet light below 2600 Å emits an intensive blue light, has found practical application. In this case the fluorescence is observed to increase in intensity with the purity of the substance, which is in favour

of the assumption that luminescence is due to the pure substance, although it should be remembered in this connection that many "Lenard luminophors" exhibit their optimum luminescence at one exact low value of impurity concentration, while arbitrary quantities of impurities reduce the luminescence. The emission and absorption spectra of the tungstates and molybdates are again composed of very diffuse bands. This, feature (not confined to these substances) makes the collection of quantitative data difficult, without which a deeper insight into the mechanism of luminescence of a solid is not feasible.

Luminescence of organic compounds

It can be generally stated that the majority of pure organic compounds, with the exception of the aromatic compounds, exhibit no or only very slight photoluminescence in the gaseous state.

The absorption spectra of many organic compounds, such as those of the aliphatic series, are continuous and accompanied by a number of diffuse bands towards the long-wave side. These bands indicate that there are certain "energised" (excited or activated) states in the molecules. If the energy of excitation or activation of such a state is greater than the binding energy of any bond in the molecule, the energy of the activated state may be utilised for the spontaneous separation of this bond (pre-dissociation). This transfer of energy within a molecule may in certain circumstances be promoted by electric fields, such as are due, for instance, to neighbouring molecules. This phenomenon of pre-dissociation which will occur the more readily the more complex the molecule, prevents the absorbed energy being re-emitted as fluorescent radiation. The absorbed energy can also be converted to thermal motion without any dissociation resulting (de-activation).

The most marked fluorescence among the aliphatic compounds is exhibited by the aldehydes and ketones (the carbonyl group $\overset{|}{\underset{|}{C}} = O$ appears to play an important part in this reaction).

Acetone vapour at atmospheric pressure consumes for pre-dissociation 17 per cent of the absorbed light quanta which are irradiated in the wave-length region of the discrete absorption bands, and only 3 per cent is emitted again as fluorescence. The remaining 80 per cent of the irradiated quanta are directly converted into heat.

Chemical reactions, also, can be induced by irradiation and this will also reduce the yield of fluorescence; thus with the aldehydes polymerisation accompanies fluorescence.

The number of aromatic compounds exhibiting fluorescence is very large. The benzene ring, the nucleus of all aromatic compounds, is very resistant to external influences, with the result that benzene and its derivatives possess a luminescence not only in the gaseous state but also in the liquid and solid states, as well as in a whole host of solvents.

The structure of the absorption and emission spectra of benzene closely resembles that of the diatomic iodine molecule already discussed. The absorption spectrum of benzene vapour consists of an extensive system of bands in the ultra-violet. This system of bands has been completely resolved and can be interpreted as the total sum of the series of oscillatory levels in the benzene ring belonging to different electronic states. The fluorescence spectrum, which is similarly located wholly in the ultra-violet, also contains a large number of bands and forms a continuation of the absorption spectrum in that the longest-wave absorption bands coincide with the shortest-wave fluorescence bands, as has already been seen in the case of the uranyl bands. No less than 450 small bands have been counted in the fluorescence spectrum of benzene.

As already stated above, luminescence is also exhibited by benzene when in solution, the bands in both the absorption and emission spectra being more or less indistinct depending on the solvent and the concentration, and being displaced slightly towards longer wave-lengths. The luminosity is a maximum at a certain concentration. Pure benzene in the liquid state has a weak luminosity, so that the strongest bands only are barely and very indistinctly visible. On cooling the benzene until crystallisation occurs, the bands become sharper again and the luminosity considerably increases. At the temperature of liquid air (—180 deg. C.) the definition again becomes comparable to that in the gaseous state (*fig. 2*).

Solid benzene exhibits a distinct afterglow, contrary to the same substance in the liquid and gaseous states in which the afterglow has a duration of only 10^{-7} to 10^{-8} sec. We shall return to this point again later.

The absorption and emission spectra of the simple benzene derivatives resemble those of benzene itself. The benzene ring retains its independence in behaviour, and the substituted radicles merely exercise a disturbing influence which is shown by a displacement of the absorption and emission spectra towards longer wave-lengths, although the general composition of the spectrum remains unaltered. At the same time the bands become less sharp

and the whole spectrum gradually merges into a continuous one. The luminescence of the benzene group is thus closely comparable to that of the uranyl group.

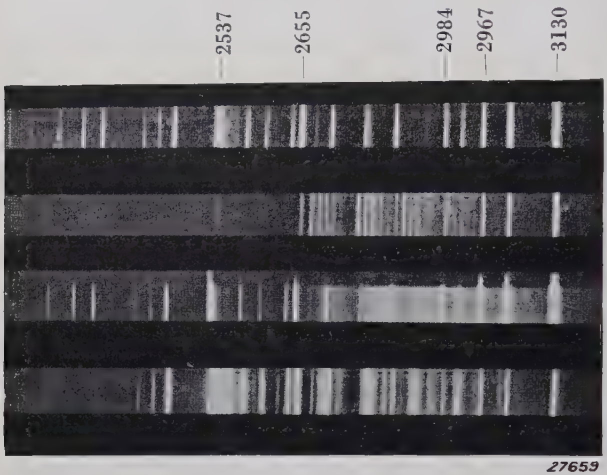

Fig. 2. Fluorescence of benzine (according to Pringsheim).
a) Mercury spectrum.
b) Fluorescence of benzene vapour.
c. Solid benzene at 0 deg. C.
d) Solid benzene at —180 deg. C.

Although very important from a theoretical point of view, the benzene derivatives are not the most striking examples of the photo-luminescence or organic compounds, since their emission spectra are situated mainly in the ultra-violet. Compounds with a visible emission spectrum have been known for many years and have been closely investigated, as for instance the highly-fluorescent solutions of certain aromatic dyes, such as fluorescein, eosin, rhodamine, etc., as well as quinine sulphate and also chlorophyll, which is the green colouring matter of plants of paramount importance in plant physiology. For these substances, neither the absorption spectrum nor the emission spectrum reveals any generalised characteristics, for both spectra depend largely on the solvent and consist of a number of usually very indistinct bands. These substances are, therefore, little suitable for an investigation and systematic co-ordination of the spectra.

Afterglow and quenching of fluorescence

The light yield of fluorescing aromatic dyes is usually a maximum at a very low concentration of the solution; as the concentration increases the yield diminishes very rapidly, which may be due to various causes. Nevertheless it has been found that in all cases this effect is the less marked the lower the mobility of the molecules of the solution, *i.e.* the more viscous the solvent. Many substances, which exhibit no or only a slight photoluminescence in liquid solvents, acquire this property to a pronounced degree in solid, vitreous solutions.

Vitreous luminescent solutions are easy to prepare. If an alcoholic solution is cooled, it becomes viscous at —125 deg. C. and at —130 deg. C. changes to a hard vitreous mass, which with a certain amount of care can be further cooled to —190 deg. C. without crystallisation occurring. Many solid solutions of this type when irradiated with a suitable wave-length are found to exhibit an intense luminescence which persists for several seconds after irradiation ceases.

This is found not only when alcohol is the solvent, but also with many other solvents. Low temperature is not a *sine qua non* for this behaviour. Solutions of many aromatic compounds in molten boric acid, which solidifies to a vitreous mass already at room temperature, also show an intense phosphorescence. In many cases the absorption and emission spectra of a substance dissolved in solid alcohol are the same as that obtained on solution in boric acid.

Various gels, such as gelatine, silica gel, albumen, cellophane, all kinds of fibrins, such as cotton wool, blotting paper, etc., can be used as adsorbents. This phenomenon is obviously again dependent on a fixation of the molecules, whereby they are protected against disturbing influences. Many substances which exhibit no luminescence whatsoever in liquid solutions fluorescence and phosphoresce strongly when they are adsorbed.

If, for instance potassium iodide is added to a solution of quinine sulphate, in sufficient quantity entirely to inhibit fluorescence when irradiated with ultra-violet light [3]), and a grain of silica gel is then dropped into the solution, the grain will show a strong fluorescence immediately it passes through an incident beam of light. By the strongly negative electrical charge on the surface of the gel, the ions $J^{-}(H_2O)$, which adversely react on the fluorescence, are kept away from the molecules of the quinine sulphate which are adsorbed by the gel.

Adsorption is probably associated with a spatial orientation of the molecules resulting in a polarisation of the light emitted. If the molecules of a dye are adsorbed by a powerful anisotropic base, such as cellophane, and if this is irradiated with non-polarised light, polarisation of the fluorescent light will occur. This orientated adsorption can probably be regarded as a transition case between the vitreous solutions and the luminophors which are produced when aromatic molecules

[3]) *v.* Philips techn. Rev., **3**, 136, 1938.

are introduced into foreign crystal lattices.

Many crystal lattices are apparently very suitable for taking up usually very small quantities of foreign molecules; examples of this are carbohydrates, bezoin acid and phthalic acid. Anthracene is also a well-known example, pure anthracene exhibiting a blue fluorescence, but ordinary commercial preparations always showing a very intense luminescence in the green (*fig. 3*), this being caused by traces of impurities, *viz.*, naphthacene, present in the lattice. Naphthacene is also an excellent example of the polarisation of light emitted by molecules occluded in crystal lattices.

In certain cases the afterglow has been closely investigated as a function of the time. In the case of dilute solutions, *e.g.* of fluorescein the duration of the afterglow is very short, and it is found that the time taken for the intensity to be reduced to one half (the half-life period) is of the order of 10^{-9} sec. Organic substances in the solid of viscous state, *e.g.* rhoduline orange in sugar solution, have an afterglow lasting several seconds. The intensity in this case can be represented with great accuracy by the exponential law:

$$I = I_0\, e^{-t/\tau}, \quad \cdot \cdot \cdot \cdot \cdot \cdot \cdot \quad (1)$$

where the constant τ is determined by the temperature, the viscosity and the concentration of the organic substance. It follows from this that we are dealing with a spontaneous phenomenon, or expressed chemically, with a mono-molecular reaction; the excited molecules return to their initial state according to a simple probability law, at the same time radiating a light quantum. The fact that the fluorescence of ordinary solutions is of such short duration and that of solid and viscous solutions is so much longer has been explained by the assumption that the excited molecule directly after absorption passes over into a metastable condition with a very low transformation probability. This problem has not been entirely cleared up.

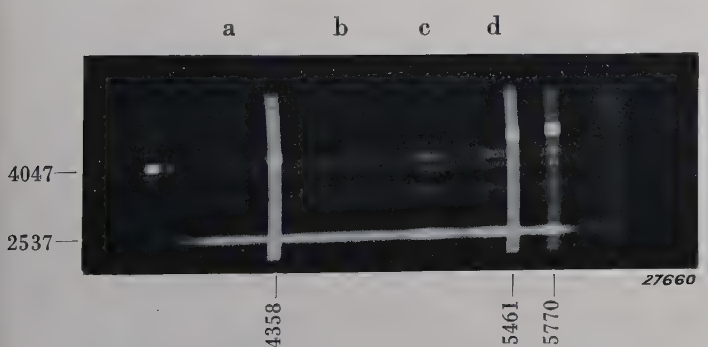

Fig. 3. Fluorescence of anthracene containing an impurity. A plate coated with anthracene was irradiated with the light from a mercury lamp decomposed into its spectrum, the long wave lengths above and the short wave lengths below. The fluorescent light emitted was resolved into a horizontal spectrum by means of a second spectrograph. Starting from the bottom the fluorescence is shown under the action of the mercury lines 2537, (2967), 3650, 4047, (4358) Å, the bracketed lines being comparatively weak. In the latter case only the bluish-green (*c*, *d*) bands are visible, and in the other cases also the blue and violet bands (*a*, *b*). The lines 5461 and 5770 Å give no fluorescence.

Inorganic "impurity" luminophors

Closely analogous to the organic "impurity" luminophors described above, are the many inorganic crystalline luminophors, which owe their luminescence to certain admixtures; a few examples of these are discussed below.

If a trace of a similar halogen compound of another metal, such as thallium, lead copper or silver, is added to an molten alkali halide, *e.g.* KCl, and the molten mass is allowed to crystallise out, a change is observed to have taken place in the absorption spectrum. The absorption spectrum of the alkali halide is composed of a number of absorption bands lying for the most part in the ultra-violet. If one of the above-mentioned metals has been added to the lattice of the alkali halide, new absorption bands make their appearance, which are characteristic for both the added metal as for the halogen. A strong luminescence is observed if light of the wave-length of these absorption bands is used for irradiation. After the termination of the period of irradiation the intensity of the afterglow decreases purely exponentially with a half-life period of about a minute.

The source of the luminescence is here a complex ion of the added metal with the halogen (*e.g.* $TlCl_4^{---}$, $PbCl_4^{--}$). which follows from the fact that the same complex ions exhibit the same fluorescence in solution. If, for instance a concentrated solution of a thallium halide is mixed with a concentrated solution of a corresponding alkali halide, the absorption spectrum of the solution is found to contain exactly the same absorption bands as the phorphorescent crystal of the alkali halide, while the solution also fluoresces.

The emission spectrum of ruby (aluminium oxide containing traces of chromium of the order of 10^{-3} per cent) consists of a large number of lines in the red, some of which occur also as absorption lines and exhibit an afterglow period of 10^{-2} to 10^{-3} sec. These lines are found to be transitions between the known energy levels of the trivalent chromium ion. In the free ion this transition would be prohibited, yet under the action of the electric fields of the neighbouring ions a finite transition probability is obtained, an interpretation which is confirmed by an investigation of the magnetic resolution of the lines.

Since the luminescence as well as the absorption of ruby is determined by the angle between the direction of the incident ray and the axes of the crystal, it is very probable that the chromium ion here forms a part of the crystal lattice and is not located haphazardly within the parent material.

Zinc sulphide exhibits a weak luminescence already in the pure state, so that with some reservation this substance can be included among those luminescent pure bodies of the type discussed above. The very intense luminescence of zinc-sulphide luminophors is however due to certain metals which are admixed with the zinc sulphide, frequently in extremely small proportions. (A concentration of 10^{-6} parts of copper is sufficient to produce a marked luminescence).

With these luminophors various phenomena have been observed which are encountered in none of the substances discussed above. In all examples dealt with up to this point the luminescence of an impurity is only observed when the wavelength of the incident light lies within the absorption spectrum characteristic of the impurity in question. But in the case of the sinc-sulphide luminophors, light which the zinc sulphide crystal itself absorbs is converted largely into a radiation which is characteristic for the sporadically-occurring impurity.

This remarkable behaviour to which the pronounced luminosity of these luminophors is due is accompanied by a second phonomenon, which has also not been observed with the substances discussed above. The zinc-sulphide luminophors are perfect insulators in the dark, but when irradiated with light which they can absorb, they become electrical conductors (photo-conductivity).

The connection between these two properties can be correlated on the following lines: During absorption of a light quantum by the zinc sulphide lattice, which is built up of a divalent positive ion Zn^{++} and a divalent negative ion S^{--}, an electron is removed from an S^{--} ion leaving an S^{-} ion; this electron attaches itself to a Zn^{++} ion which is thus converted into a Zn^{+} ion. The electron does not, however, always remain attached to the same zinc ion, but may migrate from one ion to another. Similarly a sulphur ion which has lost an electron may take up another electron from one of the neighbouring sulphur ions, which in its turn is converted to an S^{-} ion. The gap formed by the absence of an electron will thus traverse the lattice in the same way as the electron itself. These two factors are responsible for the electrical conductivity.

When by recombination an electron again fills a gap, the absorbed energy is again liberated and it is found that this energy is converted into radiation with a high quantum yield, which is characteristic of the activating metal sporadically present. The exact mechanism of this transformation of energy is still unknown, but tentatively the best explanation which can be advanced is that the metal impurity acts as a catalyst in the recombination of the electron to fill the ionic gap, as a result of which a part of the energy of recombination is transmitted as activation energy to the metallic impurity and again emitted as radiation by this component.

This explanation receives strong support from a third phenomenon, which again has not been observed with any of the luminophors previously discussed. The intensity of the afterglow when plotted as a function of the time (*fig. 4*) is not an

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Fig. 4. Photograph of a revolving disc coated with an ZnS-Cu luminophor. A mercury spectrum is projected on one radius of the disc, the longer wave lengths giving complete circles (long afterglow), and the shorter waves incomplete circles (short afterglow); the rotation time of the disc was of the order of several seconds.

exponential curve, at least during the first few seconds, but approximates more to a "hyperbolic" law:

$$I = \frac{I_0}{(1 + t/\vartheta)^2} \quad \ldots \ldots \quad (2)$$

In physical chemistry, this behaviour is recognised as a typical time function of a bimolecular reaction. If two types of atoms, A and B, are present in the same concentration, and a reaction:

$$A + B \rightarrow AB$$

takes place, then the reduction in concentration n with time is given by the expression:

$$-\frac{\mathrm{d}n}{\mathrm{d}t} = bn^2. \quad \cdots\cdots \quad (3)$$

The solution of this differential equation is:

$$n = \frac{n_0}{1 + bn_0\, t}. \quad \cdots\cdots \quad (4)$$

If A is an electron and B a gap due to the absence of an electron, and if the filling of the gap by A causes the emission of a light quantum, then if this combination takes place according to the law of bimolecular reactions, the luminous intensity is given by:

$$I = -\frac{\mathrm{d}n}{\mathrm{d}t} = \frac{b\, n_0^2}{(1 + bn_0 t)^2}. \quad \cdots \quad (5)$$

This expression is identical with (2) on substituting:

$$I_0 = bn_0^2, \qquad \vartheta = 1/bn_0 = 1/\sqrt{bI_0}.$$

The very long afterglow observed with many zinc sulphides on irradiation with light of suitable wave length is very probably due to a more complex mechanism.

In conclusion, reference must be made to a few other cases in which the luminescence is similarly due to traces of foreign metals introduced into a crystal lattice. Thus calcium tungstate, which has already been referred to as a luminescing substance in the pure state, can be caused to give a luminescence by adding *e.g.* traces of samarium, which contains in addition to the (blue) tungsten spectrum also the (red) samarium spectrum.

Also many silicates, which are not luminescent in the pure state, luminesce to a marked degree when traces of manganese are added. Both the activated tungstates and the luminescent silicates have many technical applications.

Another example of a line and band fluorescence due to the presence of traces of impurities is the mineral fluorite, which in the pure state (CaF_2) is not luminescent. Traces of the rare earths, *e.g.* europium, produce highly-fluorescent crystals which are widely found in Nature and which can also be prepared in the laboratory (*fig. 5*). Whether

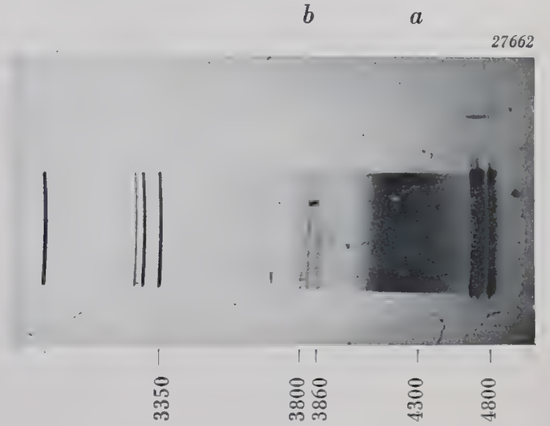

Fig. 5. Fluorescence spectrogram of a blue-fluorescing fluorite on irradiation with light from a zinc lamp (2138 Å). The sharp spectral lines are other zinc lines which also are present, as the light used for irradiation was not sufficiently monochromatic. *a* - blue bands (Eu), *b* - line fluorescence (Tb?).

this luminescence is due to the alkali halides or to the zinc sulphides has not yet been definitely established.

APPLICATIONS OF CATHODE-RAY TUBES. III.

by H. VAN SUCHTELEN.

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Investigation of high-frequency phenomena

In the previous article of this series, reference has already been made to the registration of a high-frequency oscillation of 470 kc/s [1]) with the oscillograph. The cathode-ray oscillograph is, of course, the most suitable apparatus for the direct investigation of high-frequency phenomena, and a number of further examples of its application in this direction are described below.

Measurement of the Depth of Modulation

The shape of the curve of the individual period of a high-frequency oscillation is usually not as interesting as with a low-frequency oscillation. High-frequency oscillations are generated and handled primarily in resonance circuits, so that a pure sinusoidal form is obtained in nearly all cases. In radio technology, however, considerable interest attaches to the behaviour of the oscillation over an interval of time which is large as compared with the natural period of the oscillation. The high-frequency oscillations encountered in practice are nearly always more or less modulated with respect to period, and the form of the modulated oscillation can be rendered directly visible with the cathode-ray oscillograph.

In general the frequency of modulation is very much smaller than the frequency of the oscillation itself, so that to investigate a complete modulation period the image studied must contain a very large number of periods of the high-frequency oscillation.

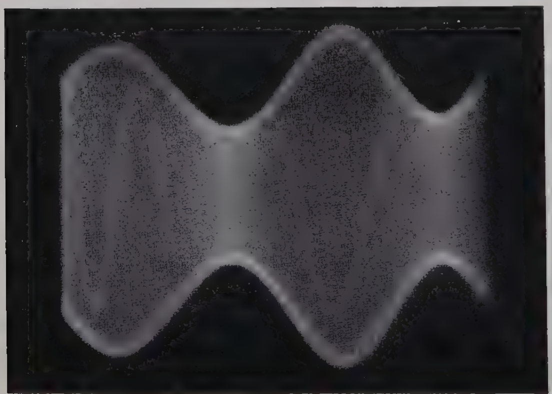

Fig. 1. Oscillogram of a 300-kc/s carrier wave modulated sinusoidally with a 400-c/s frequency. The modulation depth is about 35 per cent. With the sawtooth frequency of 200 c/s used here, the waves of the high-frequency oscillation can no longer be distinguished separately, although owing to the very rapid flyback a few high-frequency oscillations can be clearly picked out.

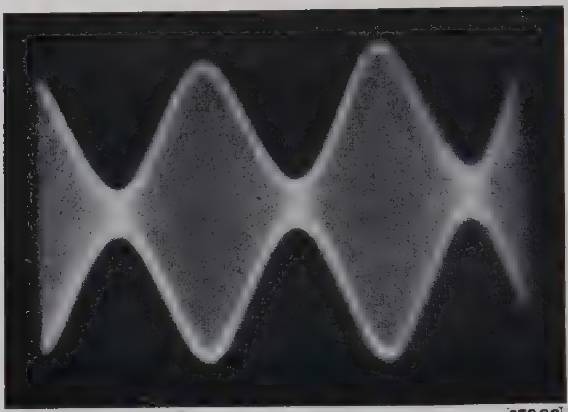

Fig. 2. Oscillogram of the same carrier wave as in fig. 1, but with a modulation depth of 75 per cent.

As a rule this number is so large that the individual curves can no longer be distinguished separately and a uniformly illuminated surface only is obtained, as shown in the registration of a sinusoidally modulated high-frequency reproduced in *fig. 1.* That the high-frequency oscillation cannot itself be differentiated is, in fact, not a disadvantage in most cases since only the envelope of the trace is of interest. A very important magnitude, the depth of modulation, can thus be read off very accurately from fig. 1, and in the case under consideration was 35 per cent, which signifies that the variation in the high-frequency amplitude was 35 per cent of the mean amplitude. In *fig. 2* an oscillogram is reproduced with a depth of modulation of 75 per cent.

In both cases the horizontal motion of the light spot was produced by a sawtooth voltage varying linearly with time [2]). To obtain a steady image the period of this time-base voltage must be equal to an integral multiple of the modulation period, and in figs. 1 and 2 this multiple was two and three respectively.

Not only is the depth of modulation obtained in this way, but also the complete geometrical form of modulation, showing, for instance, whether it is

[1]) Philips techn. Rev., **3,** 150, fig. 6, 1938. Another direct oscillogram of high-frequency oscillations is included in an article by L. Blok dealing with high-frequency oscillations in sodium lamps, Philips techn. Rev., **1,** 87, 1936. Fig. 2.

[2]) Regarding the generation of this voltage in oscillographs, *cf. e.g.* Philips techn. Rev., **1,** 147, 1936.

sinusoidal or not. It is, of course, assumed that in all cases modulation is with a sound of constant intensity, as is frequently the case in laboratory measurements.

Other conditions exist during measurements when working on radio transmitters, for the high-frequency oscillation is then usually modulated by speech or music. If an attempt is made to obtain an oscillogram of the type shown in figs. 1 or 2, the image will be found subject to such continuous and rapid changes that it can no longer be steadied by a suitable choice of the time-base frequency. Nevertheless, it is still possible to obtain an indication of the instantaneous depths of modulations in this case also, *viz.*, by eliminating the modulation frequency; this is done by employing for the horizontal deflecting voltage the low-frequency alternating voltage used for modulating the high-frequency oscillation in the transmitter, instead of a sawtooth voltage.

a

b

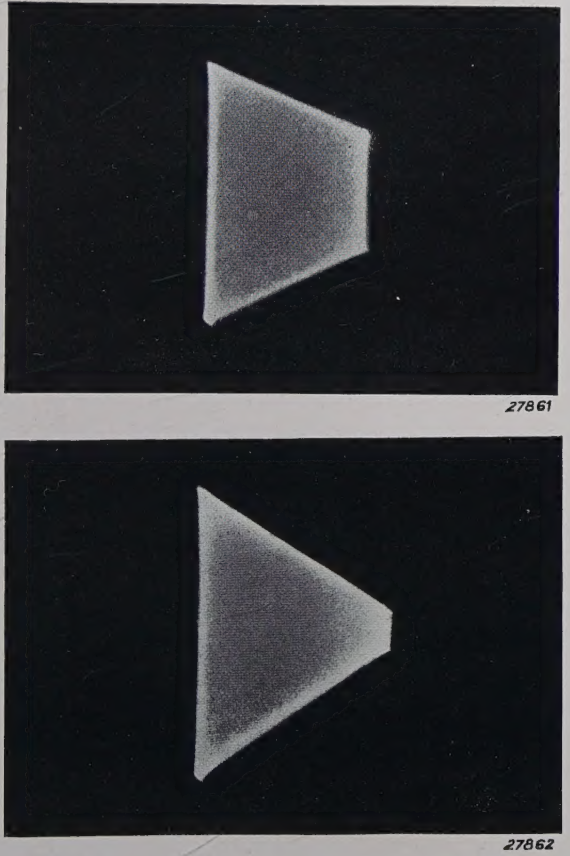

Fig. 3. Oscillogram of the same modulated high-frequency signal as in figs. 1 and 2, but plotted as a function of the instantaneous value of the modulation voltage instead of time. *a*) 35 per cent modulation, as in fig. 1; *b*) 75 per cent modulation, as in fig. 2.

The high-frequency amplitude is then no longer reproduced in the oscillogram as a function of the time, but as a function of the instantaneous value of the modulated voltage. A diagram obtained by this method is shown in *figs. 3a* and *3b*, which correspond to figs. 1 and 2. In these oscillograms the depth of modulation may be deduced from the lengths of the two vertical bounding lines, while the two sloping lines indicate if any distortion occurs in the modulating stage of the transmitter. A pure linear relationship between the high-frequency amplitude and the modulation voltage is naturally desired.

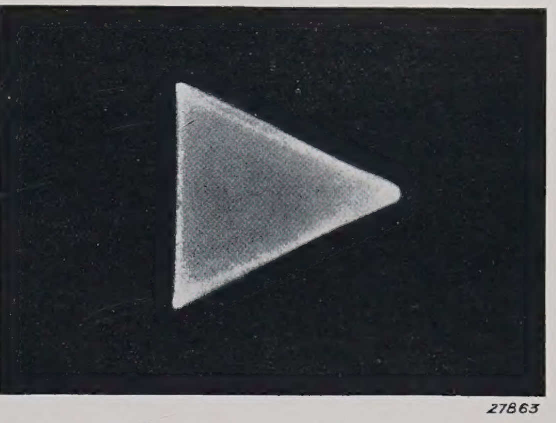

Fig. 4. Oscillogram of the same carrier wave as in figs. 3*a* and *b*, but with a depth of modulation of 95 per cent.

If modulation is not of constant depth as in figs. 3a and 3b, but is produced by speech and music, fig. 3 will pass over into *fig. 4* or *vice versa*, the change being in fact extremely rapid. The appearance of the diagram is determined by the boundaries, which correspond to the maximum depth of modulation obtained. The sloping sides of the various trapezia will always form a single straight line or a curve, while the vertical sides move rapidly to and fro. A triangle is obtained at the moment when the depth of modulation reaches the theoretical maximum value of 100 per cent (fig. 4).

The depth of modulation can also be indirectly measured in various ways and read off on an indicator, but owing to the inertia of the instruments used it is not possible to determine the very short-period maxima which may occur if the depth of modulation is not limited. The method described above, employing a cathode-ray oscillograph, can be used in this case with excellent results. The maximum depth of modulation occurring during a part of the sending interval can in fact be quite easily recorded photographically by this means, and such records may prove useful in the operation of radio transmitters.

Investigation of frequency modulation

In the above method of amplitude modulation the amplitude of the oscillation is altered with

respect to time. The frequency also can be made to vary with regard to time; this method of modulation is termed frequency modulation, and can be used for the transmission of speech and music in the same way as amplitude modulation.

In many cases, frequency modulation appears in the form of an undesirable disturbance as a result of the inadequate smoothing of the feed voltages for the amplifying valves. The modulation frequency is then as a rule equal to the mains frequency or a simple multiple thereof.

It is a comparatively simple matter to measure a frequency modulation with the cathode-ray oscillograph provided it is not too small. If the high-frequency is oscillographed over a period of the modulation frequency, the frequency of the oscillation cannot be obtained by direct computation at every point of the oscillogram as the latter is traced much too compactly, and since the frequency of the carrier wave has an altogether different order of magnitude to that of this mains voltage. However, the heterodyne method, which is extensively used in radio circuits, offers a comparatively simple means for reducing or increasing a given frequency in an arbitrary constant ratio [3]), thus if two alternating voltages are applied simultaneously to the grid of an amplifying valve having a curved characteristic, the anode current will be found to contain, in addition to the original frequencies, also the additive and differential frequencies.

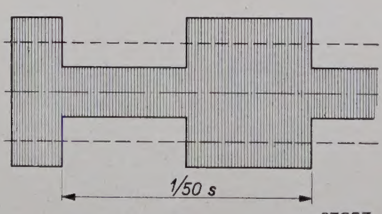

Fig. 5. Oscillogram of a carrier wave of 10^7 c/s with rectangular modulation of 50 c/s. This modulation is particularly suitable for investigating whether the carrier frequency depends on the modulation voltage (see fig. 6).

If, for instance, the frequency modulation of a high-frequency current of 10^7 c/s is to be studied, a mean differential frequency of $\nu = 1000$ c/s can be obtained by superposing an oscillation with a constant frequency of $10^7 + \nu$. This differential frequency will then undergo the same absolute variation as the initial oscillation of 10^7 c/s. If the differential frequency is oscillographed over the interval of a modulation period, $^1/_{50}$ sec., an average of $\nu : 50$ oscillations can be counted over this interval, while if ν is a frequency of 500 to 1000 c/s, these oscillations can quite easily be picked out and the variations during a modulation period can be determined directly.

As an example, an oscillogram is reproduced in *fig. 6* which was obtained in the following way: A high frequency oscillation of 10^7 cycles was modulated by a 50-cycle frequency in such a way that its amplitude had the rectangular form shown in *fig. 5*. Owing to a certain lack of efficiency of the circuit used, a frequency modulation was also obtained.

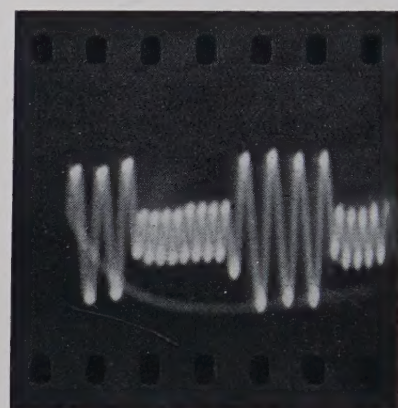

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Fig. 6. Oscillogram of the voltage obtained by applying the heterodyne principle to the carrier in fig. 5 and an auxiliary voltage of 10^7 + approximately 500 c/s. The differential frequency increases slightly with diminishing modulation voltage.

The oscillation to be investigated together with the signal of an auxiliary generator was passed to a mixing valve, and the oscillation in the anode circuit passed through a suitable filter to an oscillograph. The time-base voltage of the oscillograph was of the sawtooth type and had a frequency of 25 c/s, which was synchronised with half the frequency of the modulation voltage of 50 c/s. The frequency of the auxiliary generator was then adjusted so that a steady image (fig. 6) of two modulation periods was obtained.

The amplitude modulation is quite distinct in this oscillogram, and at the lower amplitude eight successive high-frequency oscillations and at the high amplitude four can be counted. As these figures were obtained over half a modulation period *i.e.* over $^1/_{100}$ of a second, the frequencies were 800 and 400 c/s respectively. The deviation from the mean value was thus plus or minus 200 c/s. The same deviation was also found in the case of the original oscillation of 10^7 c/s, thus showing that a very small percentage of frequency modulation can be determined in this way. In the case of sinusoidal instead of rectangular modulation a similar

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[3]) Regarding the heterodyne principle, *cf.* Philips techn. Rev., **1**, 76, 1936.

alteration is naturally more difficult to count up in the oscillogram, but it is still quite distinct.

"Blocking" of oscillators

Some oscillograms obtained in the investigation of "blocking" of a back-coupled oscillating valve are given as a third example of the application of the cathode-ray oscillograph in high-frequency work. The principle of back-coupling is shown in *fig. 7*. An oscillation can be maintained in the tuning circuit LC by introducing a sufficiently tight coupling between the coils M and L which compensates the losses in the oscillating circuit from the anode circuit. The alternating voltage applied to C produces an alternating voltage in the anode circuit by reacting on the control grid, and at the same time, by rectification in the grid circuit, the condenser C_g receives a charge which is roughly proportional to the amplitude of the oscillation and which makes the grid negative with respect to the cathode.

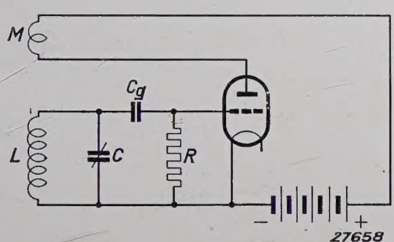

Fig. 7. Principle of an oscillating circuit.

The higher the negative bias on the grid the lower will be the amplification of the valve, and hence the smaller the energy passed to the grid circuit from the anode circuit. In ordinary circumstances an equilibrium will be reached in which the amplitude of the oscillation and the negative bias of the grid are so high that the losses in the tuning circuit are just compensated.

If the coupling between M and L is too tight, the equilibrium may be exceeded and as a result the negative bias of the grid will become so great that oscillation ceases. The negative charge will then be dissipated through R in *fig. 8* and the circuit will again start oscillating. This process is usually repeated with a frequency lying in the audible range and becomes apparent as a crackling in the loudspeaker of the receiver.

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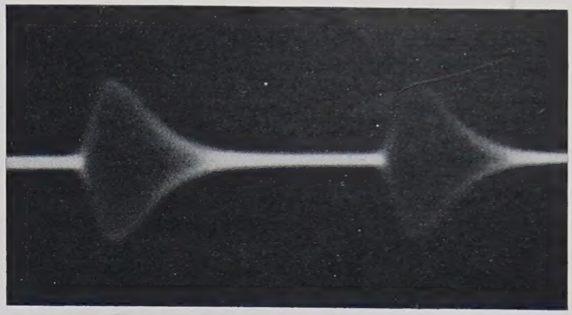

Fig. 8. "Blocking" of an oscillating circuit. If the amplitude of the oscillations exceeds a certain critical value, the valve suddenly ceases to oscillate. The oscillations then start decaying exponentially. Only when the oscillations have nearly become zero does the valve start oscillating again.

An oscillogram is shown in fig. 8 of the voltage in the tuning circuit when these conditions obtain, the tuning circuit being tuned to 1000 kc/s. It is seen that oscillation at first increases very rapidly and then beyond the critical limit decays exponentially.

In certain cases, a second tuning circuit of similar tuning characteristics may be coupled to the oscillating circuit. If the coupling between the two circuits exceeds a certain critical value, part of the oscillation energy will flow to and fro between the two oscillating circuits and as a result the voltage will exhibit fluctuations of the type shown in *fig. 9* for one of the oscillating circuits.

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Fig. 9. "Blocking" of an oscillating circuit with a second circuit coupled to it. During the repeated decay of the oscillation voltage, fluctuations occur since a part of the oscillating energy passes periodically from one oscillating circuit to the other and *vice versa*.

ABSTRACTS OF RECENT SCIENTIFIC PUBLICATIONS OF THE N.V. PHILIPS GLOEILAMPENFABRIEKEN

1294*: H. Bruining: On the emission of secondary electrons by solid substances (dissertation, Leiden, April 1938, 119 p.).

The writer contributed an article on this subject in the March number of this periodical (Philips techn. Rev. **3**, 80, 1938).

1295: C. J. Dippel and J. H. de Boer: Interlamellare Quellung von Vakuumsublimierten CaF_2-Schichten durch adsorbiertes Caesium (Rec. Trav. chim. Pays Bas **57**, 277 - 290, Mar. 1938).

The ratio is determined between caesium and iodine molecules adsorbed on layers of $Ca\ F_2$ which have been deposited by sublimation in a vacuum and then sintered. Such sintered layers have fewer capillaries and interlamellary spaces, because the adsorption of iodine is much less intense than on layers sublimed but not sintered. Caesium however is adsorbed by both surfaces to the same degree, and the adsorption only becomes less when the surface has been heated for several hours at 450° C. The different behaviour of the adsorbing surface with respect to iodine and caesium cannot apparently be explained as a poisoning by adsorption of water molecules. It is shown that a sintered surface upon which caesium has been adsorbed, adsorbs just as much iodine after the caesium has been removed as before sintering. Caesium is therefore able, by means of an interlamellary swelling, to bring the surface back to its original size without reducing the dimensions of the crystals.

1296: J. E. de Graaf: The diagnosis of casting faults with the help of X-rays (Gieterij **12**, 31 - 35, Mar. 1938).

After an introduction on the X-ray examination of macro structures (*cf.* Philips techn. Rev. **2**, 315, 1937), the diagnosis of casting faults is discussed (*cf.* Philips techn. Rev. **2**, 377, 1937). Special attention is given to the chief faults, several characteristic forms of which are treated.

*) An adequate number of reprints for the purpose of distribution is not available of those publications marked with an asterisk. Reprints of other publications may be obtained on application to the Natuurkundig Laboratorium, N.V. Philips' Gloeilampenfabrieken, Eindhoven (Holland), Kastanjelaan.

1297: J. L. Snoek: Kristallorientierung und interkristalline Korrosion (Z. Metallk. **30**, 94, Mar. 1938).

The boundary between two crystals of nickel iron with almost the same orientation oxidizes less easily than is the case with greater differences in orientation. The importance of this phenomenon is pointed out in connection with intercrystalline corrosion.

1298: K. F. Niessen: Über das Feld einer vertikalen Halbwellenantenne in beliebiger Höhe oberhalb einer ebenen Erde beliebiger Konstanten (Ann. Physik **31**, 522 - 539, März 1938).

The Hertz vector is calculated for a vertical half-wave aerial situated above a plane earth with an arbitrary dielectric constant and conductivity. This calculation is carried out to a degree of approximation such that it is permissible to use the formula for a much smaller distance of the transmitter than that to which the reflection formula applies for the case of dipole radiation .The method may be equally well used for an aerial of quite other than half wave length.

1299: J. van Niekerk and F. Franken: Hypervitaminosis-D durch grosse Gaben tierischen, antirachitischen Vitamins an Küken. (Abt. brevia Neerl. **8**, 13 - 15, Febr. 1938).

In this article the influence is discussed of excessive doses of antirachitic vitamine of animal origin on the growth and death rate of chicks.

In June 1938 appeared:

Philips Transmitting News 5, No. 2:

C. G. A. von Lindern and G. de Vries: An ultra-short telephone link between Eindhoven and Tilburg.

Four shortwave broadcast transmitters, type KVFH 10/12a, for British India.

Tj. Douma, Internal inductance of coils and its influence on the temperature coefficient of the coil.